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MASTERING PHYSICS

**With worked examples
and questions and answers**

**H.J.P.KEIGHLEY F.R.McKIM
A.CLARK M.J.HARRISON**

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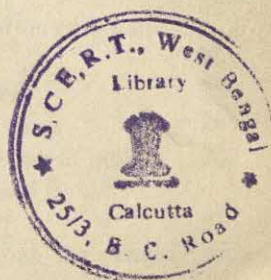
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MASTERING PHYSICS

REVISED EDITION

H. J. P. KEIGHLEY
F. R. McKIM
A. CLARK
M. J. HARRISON

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MACMILLAN



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PREFACE TO THE

FIRST EDITION

The book aims to provide a concise, easily readable treatment of all the essential principles contained in O-level physics courses. We have tried to present them with a directness and simplicity that will enable students to achieve maximum comprehension in the shortest possible time. Many diagrams have been included in the text as these are a great help in understanding physics and are especially useful in revision.

Important laws and definitions are set in italic type.

SI units have been used throughout.

Marlborough, 1982

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PREFACE TO THE REVISED EDITION

In this edition we have added some worked examples and provided answers to the questions at the end of the book. We have also extended one or two sections to provide more comprehensive coverage of the many different O-level syllabuses, and have made a few minor changes in the text to improve its clarity.

Marlborough, 1983

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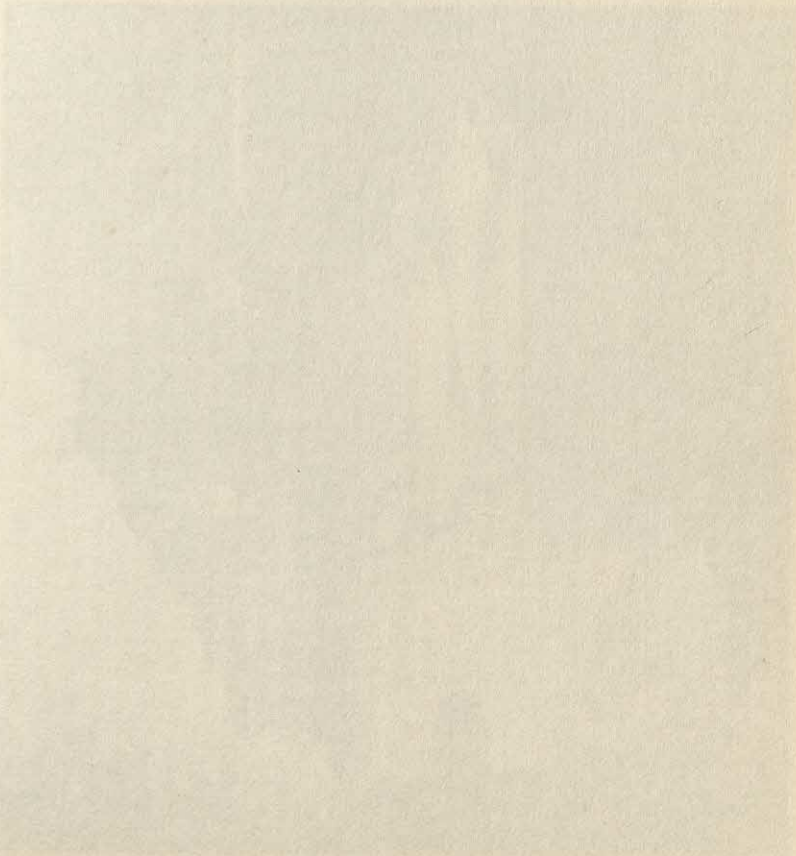
We gratefully acknowledge the following for supplying photographs which we have used to illustrate the part-title pages of the book:

Part I The Columbia space shuttle being launched at Cape Canaveral (courtesy Space Frontiers Ltd).

Part II The European Space Agency's European communications satellite (courtesy British Aerospace).

Part III Approach and runway lights as seen from the flight deck of an aircraft (courtesy British Aerospace).

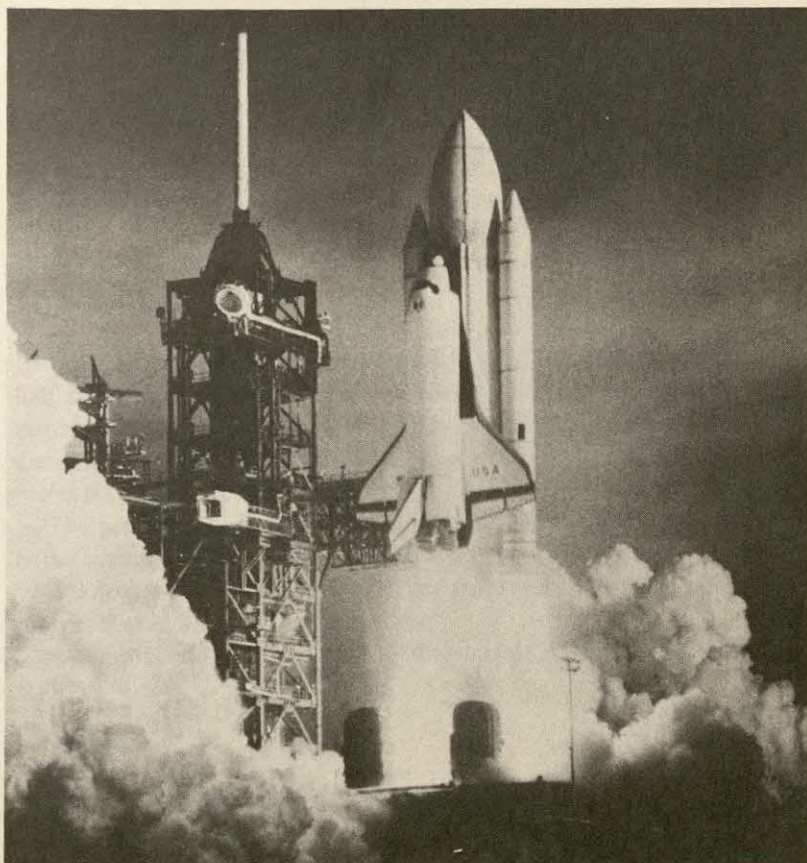
Part IV The lights of Piccadilly Circus (courtesy United Kingdom Atomic Energy Authority).



PART I

MECHANICS

The Columbia space shuttle being launched at Cape Canaveral (courtesy Space Frontiers Ltd)



FUNDAMENTAL CONCEPTS

If a study of physics or any other science is made in any detail, it is necessary to understand certain basic ideas and to know the meanings of the words used to express them. In this first chapter, we deal with some of them.

1.1 LENGTH

The standard of length is the metre (m). The marks and subdivisions on any ruler are made by comparison with the standard metre.

1.2 METRIC PREFIXES

In order to express the distance travelled in a motor car in the course of half an hour, it is usually more convenient to talk in terms of kilometres rather than metres (1 kilometre is 1000 metres). When measuring smaller distances, it is often convenient to measure them in millimetres (1 millimetre is one-thousandth of a metre).

The prefixes in common use with metric units are shown in Table 1.1. Some examples of the use of these prefixes are as follows:

A kilogram is 10^3 grams, that is, 1000 grams; a megatonne is 10^6 tonnes, that is, a million tonnes.

A microsecond is 10^{-6} second, that is, one-millionth of a second.

1.3 FORCE

The boy shown in Fig. 1.1a is being pushed and in Fig. 1.1b is being pulled. Whenever you pull or push something, you are applying a force to it. A force may

- (i) change the speed of an object;
- (ii) change the shape or size of an object;
- (iii) change the direction in which an object is travelling.

Can you think of examples to illustrate each of the above statements?

Table 1.1 *prefixes used with metric units*

<i>Submultiple</i>	<i>Prefix</i>	<i>Symbol</i>
10^{12}	tera-	T
10^9	giga-	G
10^6	mega-	M
10^3	kilo-	k
10^1	deca-	da
10^{-2}	centi-	c
10^{-3}	milli-	m
10^{-6}	micro-	μ
10^{-9}	nano-	n
10^{-12}	pico-	p

fig 1.1 a force may be thought of as (a) a push or (b) a pull



(a)



(b)

One force with which everyone is familiar is the force of gravity. If you let go of a stone which you were holding, the stone starts to move towards the ground because the force of gravity is pulling it downwards. While you were holding the stone, it did not move because you were pulling up on the stone with a force equal to the force of gravity downwards on the stone. In that case there were two equal and opposite forces acting on the

stone and we say that the *resultant* force was zero. When the resultant force on a body is zero, it will not change its speed nor its direction of travel.

1.4 MEASUREMENT OF FORCE

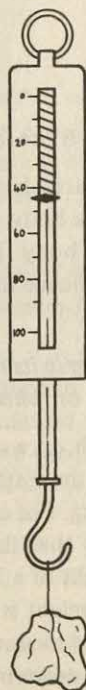
A simple way of measuring a force is to use a spring balance (Fig. 1.2). The greater the force that is applied, the greater is the movement of the pointer attached to the spring. Providing the force applied to a spring is not too large

The extension of the spring is proportional to the force

(This is known as *Hooke's law*.)

Force is measured in units called newtons and spring balances are calibrated in newtons.

fig 1.2 a spring balance may be used to measure the force on a mass hung from it



In Section 4.4 we show how this unit of measurement was chosen, but it might help for the moment if you remember that it requires a force of 1 newton (1 N) to support an average-sized apple. If an average-sized apple is hung on a spring balance, the balance will read 1 N.

1.5 MASS

As we stated above, a force can cause a body to increase its speed, that is, to accelerate. Will the acceleration of a small car pushed by three people be the same as that of a large lorry pushed just as hard by the same three people? Obviously not. Clearly the acceleration for a given force depends on the body being accelerated. We say that a body which is difficult to accelerate is more massive than a body which is easy to accelerate. The mass of a body is a measure of how difficult the body is to accelerate.

The unit of mass is the kilogram (kg)

This standard is a lump of platinum alloy kept at Sevres near Paris. If masses of 2 kg and 1 kg have the same force applied to them, then the 1 kg mass will have twice the acceleration of the 2 kg mass.

It is sometimes helpful to think of mass as the quantity of 'stuff' that is in the body. A mass of 2 kg has in it twice the quantity of 'stuff' that is in a mass of 1 kg.

1.6 WEIGHT

The weight of a body is the force of gravity on it

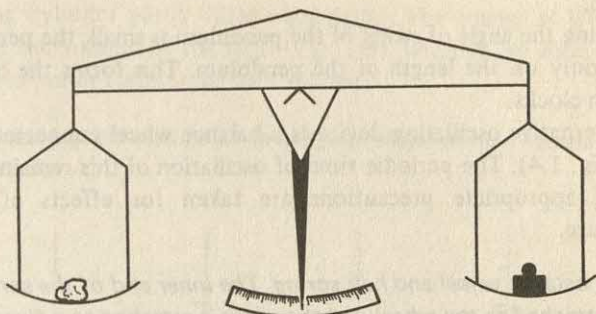
The weight of a body on the Earth is the force with which the Earth attracts the body. The weight of a body on any other planet is the force with which the planet attracts the body. The force of gravity on the Moon is less than that on the Earth and hence the weight of a body on the Moon is less than its weight on the Earth.

Wherever a body is in the universe its mass remains constant, but its weight depends on where it is situated

The force exerted by the Earth on a mass of 1 kg is 9.8 N. We sometimes express this by saying that the Earth's gravitational field is 9.8 newtons per kilogram, written 9.8 N/kg. For convenience in calculations this is often taken as 10 N/kg. It follows that the pull of the Earth on a mass of 2 kg is $10 \times 2 = 20$ N, and the weight of a 3 kg mass is $10 \times 3 = 30$ N.

We have already stated that weight is measured with a spring balance calibrated in newtons. Mass may be measured with a beam balance like the one illustrated in Fig. 1.3. The unknown mass is put in one of the pans and

fig 1.3 a beam balance compares the masses of objects in each of the two pans



known masses are put in the other pan until the arm of the balance is horizontal. The masses in each of the pans are then the same since the Earth pulls equal masses with the same force.

Other types of balance can be used to measure mass. Often only one pan is used and balance can be obtained by sliding special masses along a scale, or sometimes the balance displays the mass of the object directly on a screen.

1.7 TIME

For many scientific experiments it is essential to be able to measure time, or a change in time, accurately. For example, it is not possible to calculate the speed of a rifle bullet unless the time can be measured during which the bullet travels a measured distance. The fundamental unit of time is the second (s) defined as follows:

The second is a specified number of periods of a certain radiation from the caesium-133 atom

For all except advanced work it is accurate enough to begin from the rotation of the Earth exactly once in one day, and divide the day into 24 hours, each hour into 60 minutes and each minute into 60 seconds. The only reason why this is not exact is that not all days are of equal length. So the definition in terms of the radiation from caesium-133 atoms is more precise.

1.8 MEASURING TIME

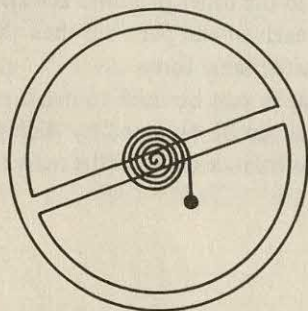
One of the most easily understood methods of measuring time is to use a pendulum. The simplest form of pendulum consists of a lump of, say,

metal on the end of a piece of string. If the string is hung from its free end and allowed to oscillate, then the time for one complete oscillation of the pendulum ('both there and back') is known as the periodic time of the pendulum.

Providing the angle of swing of the pendulum is small, the periodic time depends only on the length of the pendulum. This forms the basis of all pendulum clocks.

An alternative oscillating device is a balance wheel connected to a hair spring (Fig. 1.4). The periodic time of oscillation of this remains constant providing appropriate precautions are taken for effects of changing temperature.

fig 1.4 *a balance wheel and hair spring. The inner end of the spring is attached to the wheel and the outer is attached to a fixed frame*



Most accurate clocks these days depend for their time-keeping on the oscillatory property of a quartz crystal, when joined into an electronic circuit. This is the timing mechanism of all quartz watches.

1.9 VOLUME

It is sometimes necessary to calculate the volume of variously shaped samples. For some standard objects the values are as follows:

For a rectangular solid of length a , breadth b and height c

$$\text{Volume} = a \times b \times c \quad (1.1)$$

For a cylinder of radius r and height h

$$\text{Volume} = \pi r^2 h \quad (1.2)$$

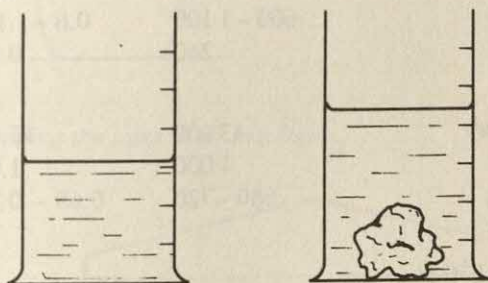
For a sphere of radius r

$$\text{Volume} = \frac{4\pi r^3}{3} \quad (1.3)$$

1.10 MEASURING VOLUME

If a solid has an irregular shape its volume can be measured by placing it in a measuring cylinder partly filled with liquid. The change in level of the liquid gives the volume of the solid (Fig. 1.5), because the solid displaces an amount of water equal to its own volume.

fig 1.5 *measuring the volume of an irregular shape using a measuring cylinder*



The volume of a quantity of liquid can be measured directly using a measuring cylinder.

1.11 DENSITY

We say that steel is heavier than wood, yet a tree is heavier than a pin. Most people understand that when we say steel is heavier than wood we are comparing *equal volumes*.

The quantity which remains constant for any sample of a particular material is the mass per unit volume. This is termed the density, so that we may write

Density is mass per unit volume

or

$$\text{Density} = \frac{\text{Mass}}{\text{Volume}} \quad (1.4)$$

In the SI system of measurement, mass has the units of kilograms, and volume has the units of cubic metres. Density is therefore in kg/m^3 .

Sometimes, partly because the numbers involved are smaller, the mass is expressed in grams (g) and the volume in cubic centimetres (cm^3), so that the density is in g/cm^3 .

The densities of some typical substances are given in Table 1.2.

Table 1.2 *densities of various substances*

	kg/m ³	g/cm ³
<i>Solids</i>		
Uranium	18 700	18.7
Copper	8 900	8.9
Aluminium	2 700	2.7
Glass (varies)	2 500	2.5
Ice	920	0.92
Wood	600 - 1 100	0.6 - 1.1
Cork	240	0.24
<i>Liquids</i>		
Mercury	13 600	13.6
Water	1 000	1.00
Petrol	680 - 720	0.68 - 0.72
<i>Gases</i>		
Air at 0°C at sea level	1.29	1.29×10^{-3}
Hydrogen at 0°C at sea level	9.0×10^{-2}	9.0×10^{-5}

A quantity called the relative density is defined by the equation

$$\text{Relative density} = \frac{\text{Density of substance}}{\text{Density of water}} \quad (1.5)$$

This is a ratio and has no units.

1.12 THE DENSITY BOTTLE

The density bottle has a ground glass stopper with a fine capillary tube running through it, along which excess liquid escapes when the stopper is inserted (Fig. 1.6). Whatever liquid is put in the bottle, exactly the same volume of liquid must fill the bottle in each case.

Below is a typical set of readings obtained using the bottle:

Mass of empty bottle	= 18.5 g
Mass of bottle full of water	= 50.5 g
Mass of bottle full of liquid	= 70.1 g

fig 1.6 a density bottle

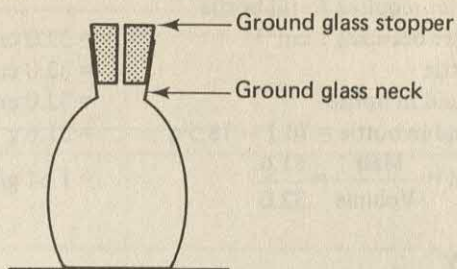
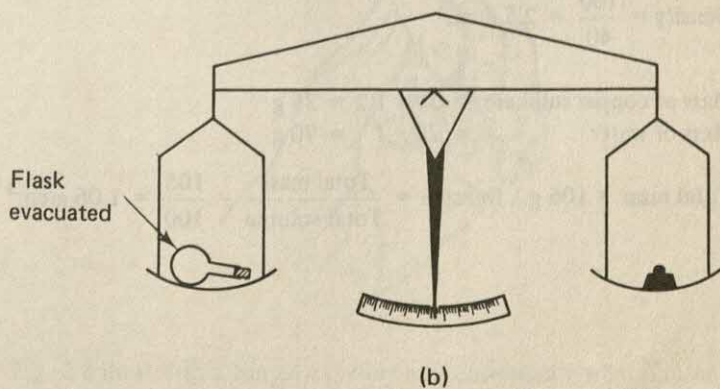
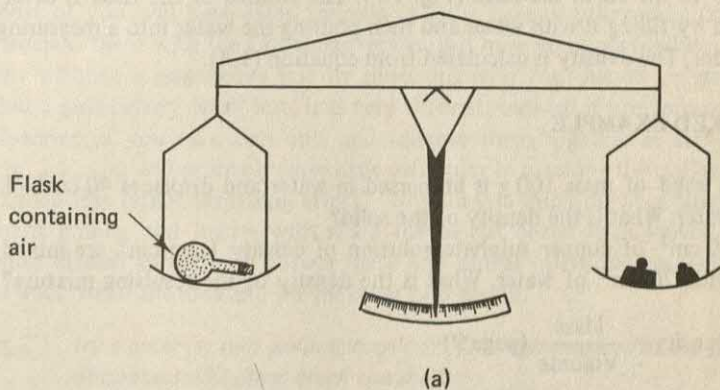


fig 1.7 determining the mass of air in a flask



The liquid density is calculated as follows:

$$\begin{aligned}
 \text{Mass of water in the bottle} &= 50.5 - 18.5 \text{ g} &= 32.0 \text{ g} \\
 \therefore \text{Volume of water required to fill bottle} && \\
 (\text{since } 1 \text{ g of water occupies } 1 \text{ cm}^3) &&= 32.0 \text{ cm}^3 \\
 \therefore \text{Volume of bottle} &&= 32.0 \text{ cm}^3 \\
 \therefore \text{Volume of liquid in bottle} &&= 32.0 \text{ cm}^3 \\
 \text{But mass of liquid in bottle} &= 70.1 - 18.5 \text{ g} &= 51.6 \text{ g} \\
 \text{Density of liquid} &= \frac{\text{Mass}}{\text{Volume}} = \frac{51.6}{32.0} &= 1.61 \text{ g/cm}^3
 \end{aligned}$$

1.13 AIR DENSITY

To determine the density of air, a half-litre flask is weighed (a) when full of air and (b) when evacuated using a vacuum pump. The difference in mass is due to the air in the flask (Fig. 1.7). The volume of the flask is determined by filling it with water and then pouring the water into a measuring cylinder. The density is calculated from equation (1.4).

WORKED EXAMPLE

- (i) A solid of mass 100 g is immersed in water and displaces 40 cm³ of water. What is the density of the solid?
- (ii) 30 cm³ of copper sulphate solution of density 1.2 g/cm³ are mixed with 70 cm³ of water. What is the density of the resulting mixture?

(i) $\text{Density} = \frac{\text{Mass}}{\text{Volume}}$ (page 9)

Since the solid displaces 40 cm³ of water its volume is 40 cm³

$$\text{Density} = \frac{100}{40} = 2.5 \text{ g/cm}^3$$

- (ii) Mass of copper sulphate = $30 \times 1.2 = 36 \text{ g}$
 Mass of water = $70 \times 1 = 70 \text{ g}$

$$\text{Total mass} = 106 \text{ g} \therefore \text{Density} = \frac{\text{Total mass}}{\text{Total volume}} = \frac{106}{100} = 1.06 \text{ g/cm}^3$$

PRESSURE

2.1 THE CONCEPT OF PRESSURE

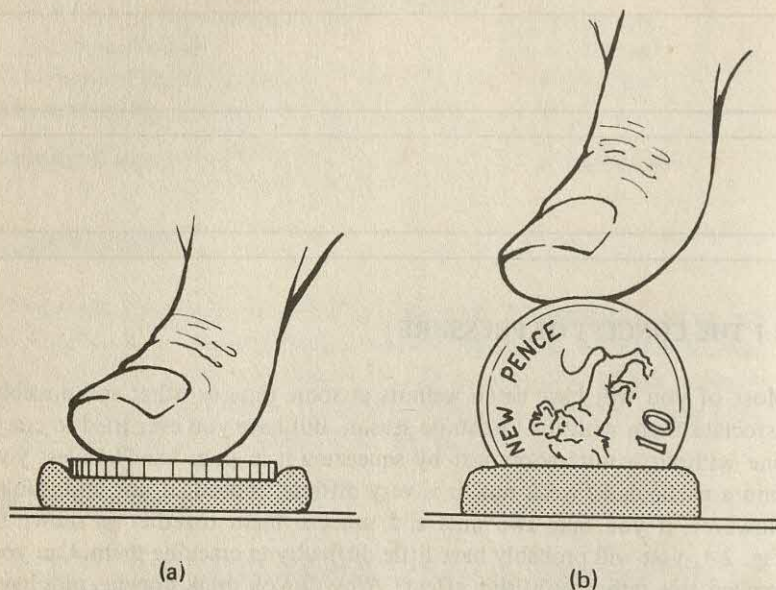
Most of you will have eaten walnuts at some time or other and possibly associate them with the Christmas season. But have you ever tried to crack one without a nutcracker just by squeezing it in your hand? Unless you find a particularly weak nut, it is very difficult indeed, if not impossible. However, if you take two nuts and squeeze them together as shown in Fig. 2.1, you will probably have little difficulty in cracking them. Can you explain this rather surprising effect? Why do you think drawing pins have sharp points, and knives with sharp blades cut better than knives with blunt blades? In order to answer these and similar questions, we must try to understand the idea and the meaning of pressure.

fig 2.1 *try squeezing two walnuts together. The high pressure at the point of contact will often crack the shells*



Fig. 2.2 illustrates a simple experiment to investigate what is meant by pressure. A coin is being pushed into plasticine first by exerting a force on

fig 2.2 the higher pressure in (b) causes the coin to sink more easily into the plasticine



the flat side of it and secondly by exerting the same force on the edge. It is obvious in which case the coin is pushed more readily into the plasticine. Many people are also familiar with the fact that if you put on a pair of skis you do not sink as far into the snow as you would do if you were standing on the snow without skis. How can we explain these facts?

How far you sink in each case is determined not only by the hardness and condition of the snow and the force exerted on it, but also by the area over which the force acts. It is the force per unit area or the *pressure*, as we call it, that determines the penetration.

Pressure is the force (or thrust) acting on unit area

i.e.

$$\text{Pressure} = \frac{\text{Force}}{\text{Area}} \quad (2.1)$$

If the force is expressed in newtons and the area in square metres, the pressure is in *pascals* (Pa), i.e. $1 \text{ Pa} = 1 \text{ N/m}^2$.

A boy weighing 600 N is standing on snow. If the shoes he is wearing have a total area in contact with the ground of 0.03 m^2 , then the pressure he is exerting on the snow is given by

$$\text{Pressure} = \frac{600}{0.03} = 20\,000 \text{ N/m}^2 = 20\,000 \text{ Pa} = 20 \text{ kPa}$$

If the boy then puts on skis of total area 0.3 m^2 , the pressure he now exerts on the snow is given by

$$\text{Pressure} = \frac{600}{0.3} = 2000 \text{ Pa} = 2 \text{ kPa}$$

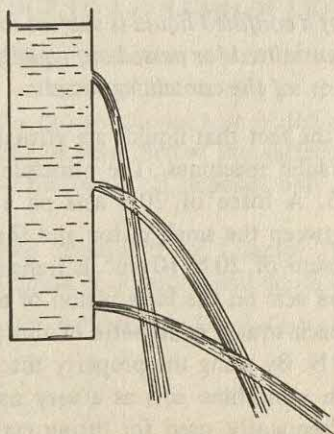
The new pressure is only one-tenth of the old pressure and this is the reason he does not sink so far into the snow.

Now think again about cracking walnuts. When two walnuts are in the hand, the area of contact between the two walnuts is quite small. Can you explain in terms of pressure why it is easier to crack a walnut if a second walnut is held against it?

2.2 PRESSURE IN LIQUIDS

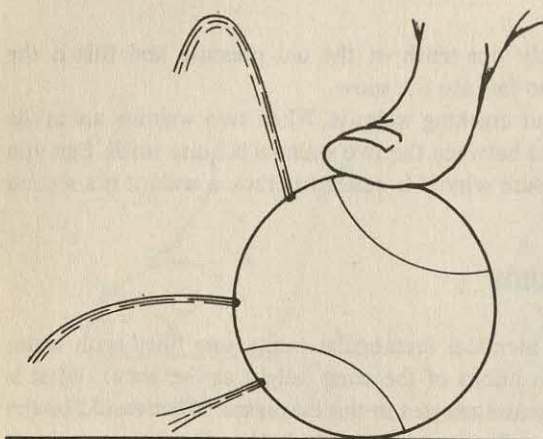
Suppose you have two identical rectangular tanks, one filled with water and the other filled with bricks of the same weight as the water. What is the difference in the pressure exerted in the two cases? What would be the effect of removing the walls in each case? In both cases the pressure on the bottom is the same but in the case of the liquid the pressure is exerted in other directions besides downwards. Indeed a liquid exerts its pressure in all directions. Fig. 2.3 illustrates a demonstration which shows the side-ways pressure of water. A vertical-sided tank has three holes drilled in the side. Notice that the water pressure at the bottom is greater than that at

fig 2.3 an experiment to show that pressure in liquids increases with depth



the top. How does the experiment show this? The fact that pressure acts in *all* directions may be indicated by the demonstration shown in Fig. 2.4. A rubber ball with lots of holes in it is filled with water. When the ball is

fig 2.4 *pressure in liquids acts in all directions*



squeezed, the water comes out of all the holes, showing that the pressure acts in all directions on the inside surface of the ball. More precise experiments show that

The pressure at any point in a liquid acts equally in all directions

The experiment with a rubber ball also illustrates another important principle, namely that

When any part of a confined liquid is subject to a pressure, the pressure is transmitted (or passed on) equally to all parts of the containing vessel

This principle and the fact that liquids are virtually incompressible are made use of in hydraulic machines. The principle of such machines is illustrated in Fig. 2.5. A force of 20 N acts on a small piston of area 10 cm^2 . The space between the small piston and the large piston is filled with liquid. The pressure of $20\text{ N}/10\text{ cm}^2$ is transmitted throughout the liquid and this pressure acts on the large piston of area 100 cm^2 . Since a force of 2 N acts on each square centimetre of this piston, the total force on it is $2 \times 100 = 200\text{ N}$. By using the property that a liquid has of transmitting pressure, such a machine acts as a very useful force multiplier. Hydraulic jacks are frequently used for lifting cars at garages. Fig. 2.6

fig 2.5 the principle of hydraulic machines

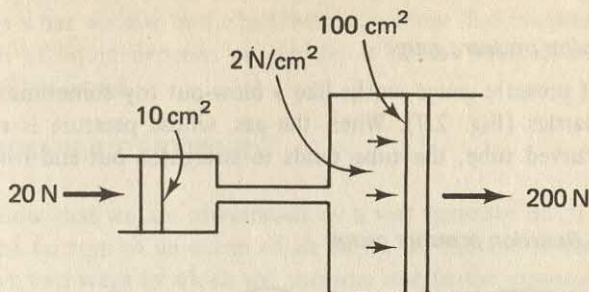
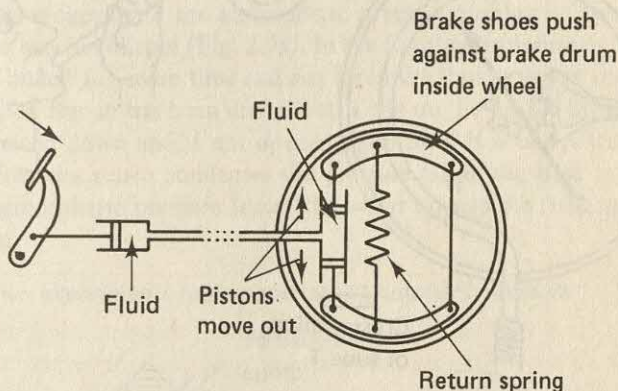


fig. 2.6 the principle of hydraulic brakes (not drawn to scale)



illustrates the principle of hydraulic brakes in cars. The pressure is transmitted through the fluid in the hydraulic system, and a small force exerted on the brake pedal can cause a large force to be exerted on the wheels.

2.3 THE PRESSURE DUE TO A COLUMN OF LIQUID

It may be shown that the pressure due to a column of liquid depends only on the height of the column and the density of the liquid. The pressure does not depend on the cross-sectional area of the column. For example, the pressure at a depth of 50 m below the surface of the sea does not depend on the size of the ocean. It depends only on the depth and the density of sea water.

It can be shown (see question 9 for Chapter 2) that the pressure (P) below the surface of a liquid of density ρ and depth h is given by

$$\text{Pressure (N/m}^2\text{)} = 10 \text{ (N/kg)} \times \text{Depth (m)} \times \text{Density (kg/m}^3\text{)} \quad (2.2)$$

or

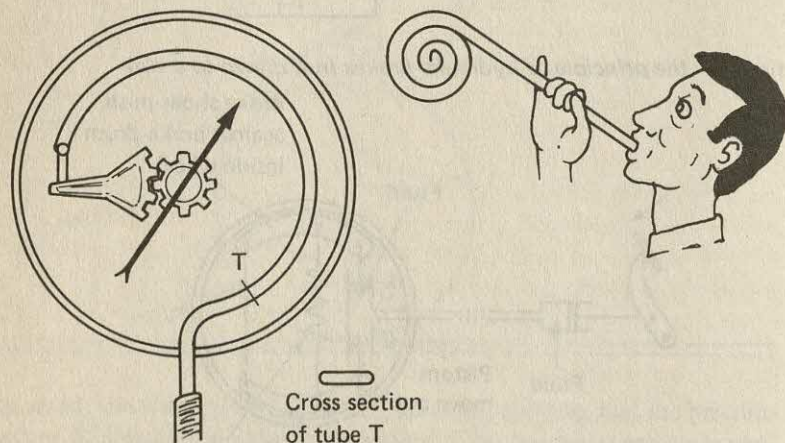
$$P = 10 h \rho \text{ N/m}^2$$

2.4 MEASURING PRESSURE

(i) *The Bourdon pressure gauge*

This type of pressure gauge works like a blow-out toy sometimes seen at Christmas parties (Fig. 2.7). When the gas, whose pressure is required, enters the curved tube, the tube tends to straighten out and rotates the pointer.

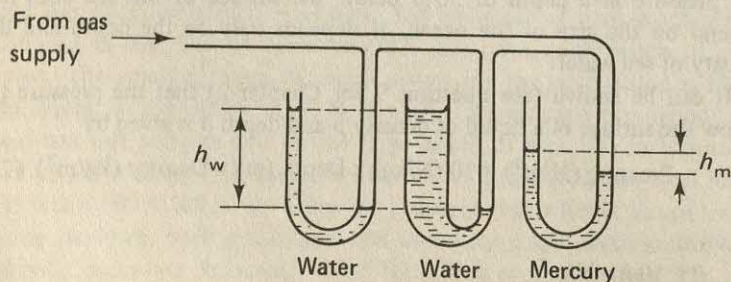
fig 2.7 the Bourdon pressure gauge



(ii) *The U-tube manometer*

Fig. 2.8 shows U-tube manometers being used to measure the pressure of the gas supply. Two of the U-tubes contain water. Notice that the height of the column of water (h_w) does *not* depend on the cross-sectional area of the tube. The height does, however, depend on what liquid is in the tube. When mercury, which is much denser than water, is used, the height of the

fig 2.8 U-tube manometers

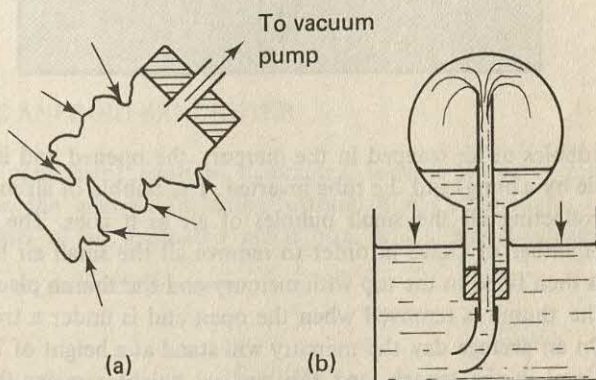


mercury (h_m) is much less than the height of the water column. This illustrates what we saw in the last section, namely that the pressure due to a column of liquid depends on the depth of the liquid column and the density of the liquid.

2.5 ATMOSPHERIC PRESSURE

We all know that we are surrounded by a vast quantity of air. In fact we live at the bottom of an ocean of air which we call the atmosphere. Fig. 2.9 shows two ways in which the pressure due to the atmosphere can be observed. In the first experiment, a metal can is connected to a vacuum pump and the air inside is withdrawn. The pressure on the inside of the can is thus reduced and the atmospheric pressure pushing on the outside causes the can to collapse (Fig. 2.9a). In the fountain experiment, water in a flask is boiled for some time and any air in the flask is driven out by the steam. When the air has been driven out, a clip on the rubber tubing at the top is screwed down and is not opened again until it is under water (Fig. 2.9b). When the steam condenses the pressure inside the flask is reduced and the atmospheric pressure forces the water up into the flask producing a fountain.

fig 2.9 *two experiments to demonstrate atmospheric pressure*



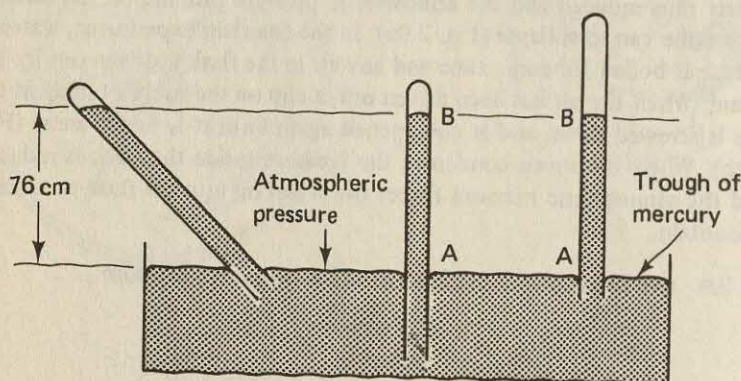
In 1651, Otto von Guericke, the mayor of Magdeburg, performed a striking experiment before the Emperor Ferdinand III to demonstrate the existence of atmospheric pressure. Having invented the first vacuum pump, von Guericke made two metal hemispheres about 60 cm in diameter which fitted together with an airtight joint. When he pumped the air out from inside the joined hemispheres, it took eight horses on each side to pull them apart. Removing the air from inside the hemispheres reduced the outward force on each hemisphere. The surrounding atmosphere still

produced an inward force which held the two hemispheres together until the forces produced by the horses were big enough to pull them apart.

2.6 A SIMPLE BAROMETER

A simple barometer may be constructed by using mercury and a glass tube, which is about 90 cm long, sealed at one end. The tube is first filled with mercury to within about 1 cm of the top (Fig. 2.10). In order to remove

fig 2.10 *the vertical height of the barometer stays constant if the tube is tilted*



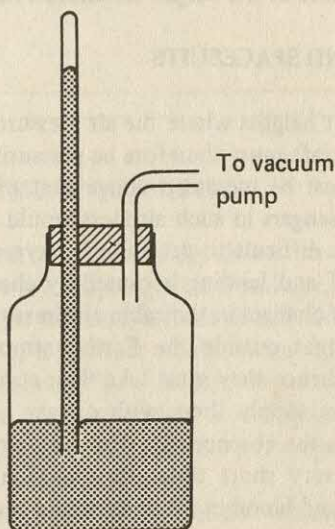
the small bubbles of air trapped in the mercury the opened end is closed (for example by a bung) and the tube inverted. The bubble of air moves up the tube collecting all the small bubbles of air as it goes. The tube is inverted a number of times in order to remove all the small air bubbles. The tube is then filled to the top with mercury and the thumb placed over the end. The thumb is removed when the open end is under a trough of mercury. On an average day the mercury will stand at a height of 760 mm above the level in the trough, and this vertical height remains the same even if the tube is tilted.

The pressure at A due to the column of mercury AB is equal to the atmospheric pressure. The vertical height AB is therefore a measure of the atmospheric pressure in millimetres of mercury (mmHg). We can calculate the pressure using equation (2.1), knowing that the density of mercury is $13\,600\text{ kg/m}^3$ and that the mercury column height is 0.76 m. Using these data

$$\text{Pressure} = 10 \times 0.76 \times 13\,600 = 1.03 \times 10^5 \text{ N/m}^2 = 103 \text{ kPa}$$

Fig. 2.11 shows an apparatus which may be used to demonstrate the fact that it is the atmospheric pressure which is pushing the mercury up the tube and not the vacuum which is sucking it up. As the air above the mercury in the trough is withdrawn, the height of the mercury in the barometer tube falls.

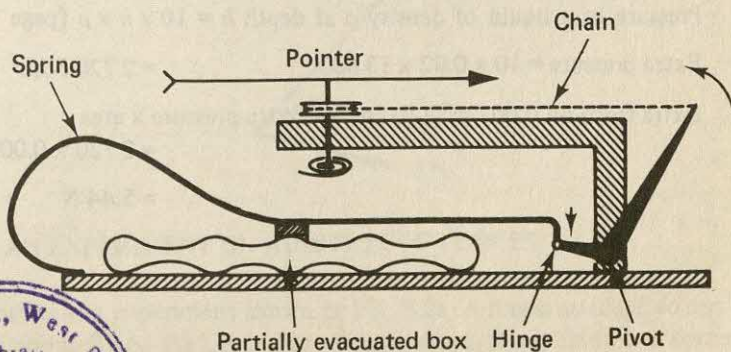
fig 2.11 if the air pressure in the flask is reduced, the height of the mercury in the tube falls



2.7 THE ANEROID BAROMETER

A very convenient form of barometer, because it can be moved about easily, is the aneroid (meaning 'without liquid') barometer (Fig. 2.12). It consists of an evacuated metal box, which has corrugated sides to

fig 2.12 an aneroid barometer



increase its strength. A fixed spring prevents the box from collapsing completely. If the atmospheric pressure increases, the centre of the box moves inwards and this movement is magnified by a system of levers. The last lever is attached to a chain which turns the pointer. A hair spring takes up any slack in the chain.

A similar instrument is frequently used as an altimeter in an aeroplane. As the aeroplane rises, the atmospheric pressure decreases, and the decrease in pressure is a measure of the height the aircraft has attained.

2.8 AIRLINERS AND SPACESUITS

Many airliners fly at heights where the air pressure is very low. The cabins of high-altitude aircraft must therefore be pressurised, that is, the pressure inside the cabin must be increased above that of the air outside. If this were not done, passengers in such airliners would experience extreme discomfort and find it difficult to get enough oxygen. The 'popping' of the ears during take-off and landing is caused by changes of pressure on the ear drum because of changes in the cabin air pressure.

Astronauts go right outside the Earth's atmosphere to places where there is no air and hence they must take their atmosphere with them. The spacesuits they wear supply them with oxygen and maintain them at a suitable pressure. In the absence of a spacesuit, a man on the Moon would be dead within a very short time. He would not have any oxygen to breathe, the water and blood in his body would boil and probably explode into the surrounding vacuum.

WORKED EXAMPLE

A piston of area 20 cm^2 fits a cylinder in which it runs smoothly. The air pressure inside the cylinder is increased by 2 cm of mercury. What extra force acts on the piston? (Density of mercury = $13\,600 \text{ kg/m}^3$; $g = 10 \text{ N/kg}$)

Pressure in a liquid of density ρ at depth $h = 10 \times h \times \rho$ (page 17)

$$\text{Extra pressure} = 10 \times 0.02 \times 13\,600 = 2\,720 \text{ N/m}^2$$

$$\begin{aligned} \text{Extra force on } 0.002 \text{ m}^2 \text{ (} 20 \text{ cm}^2 \text{)} &= \text{Extra pressure} \times \text{area} \\ &= 2\,720 \times 0.002 \text{ N} \\ &= 5.44 \text{ N} \end{aligned}$$

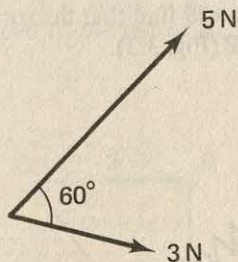
VECTORS

3.1 ADDITION OF PHYSICAL QUANTITIES

Does a mass of 3 g added to a mass of 5 g always produce a total mass of 8 g? You may think this is a stupid question and say, 'Of course it does'. You would be quite right. But does a force of 3 N added to a force of 5 N always produce a force of 8 N? What would be the result if they pulled on a body in opposite directions? Two forces, one of 5 N one way and the other of 3 N the other way are equivalent to one single force of 2 N in the first direction. The single force to which the other two added together are equivalent is termed the resultant force.

Suppose the two forces make an angle of 60° with each other, as shown in Fig. 3.1. What is the resultant force then? We can answer this question from the results of the experiment described below.

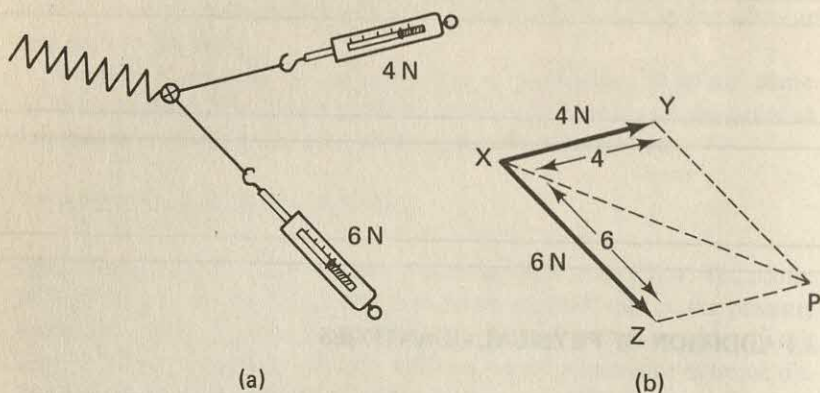
fig 3.1



3.2 AN EXPERIMENT ON ADDITION OF FORCES

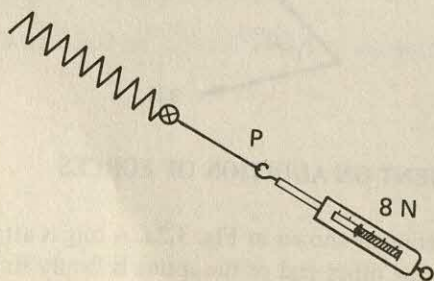
Consider the experiment shown in Fig. 3.2a. A ring is attached to one end of a spring S and the other end of the spring is firmly fixed to a corner of

fig 3.2 the diagonal XP represents the resultant of the two forces represented by the lines XY and XZ



a large board (e.g. blackboard). Two spring balances are attached to the ring at different points so that the spring S may be extended by pulling on the spring balances. When the spring is steady, the position of the centre of the ring X is marked on the board. Lines XY and XZ are drawn to show the direction in which each spring balance is pulling. The spring balances are then removed from the ring. Suppose that before they were removed the balances read 4 N and 6 N. Fig. 3.2b shows the line XY drawn 4 units long and the line XZ drawn 6 units long. These lines then represent the forces both in magnitude (size) and direction. The parallelogram XYPZ is then completed. The diagonal XP is 8 units long. You may well have guessed that this 8 units represents the resultant of the two forces of 4 N and 6 N. This may be verified by fixing a spring balance to the ring and pulling on it until the ring is in exactly the same position as in the first part of the experiment. You will find that the spring balance reads 8 N and is pulling in the direction XP (Fig. 3.3).

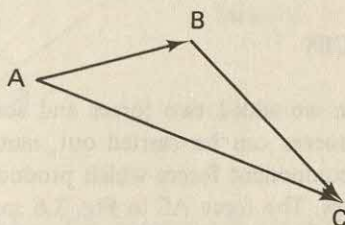
fig 3.3



It is not necessary to draw the whole parallelogram in order to add two forces. The rule for adding two forces is as follows:

Represent one of the forces in magnitude and direction by the line AB (Fig. 3.4). At the point B draw another line BC to represent the other force in magnitude and direction. The line AC represents the resultant force in both magnitude and direction

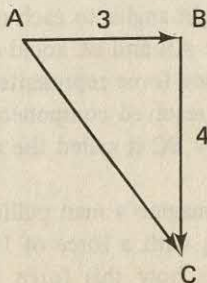
fig 3.4 *the resultant of the forces represented by AB and BC is represented by the line AC*



The same rule applies for adding velocities. Suppose you are rowing across a stream at 3 metres per second (m/s) and that the stream is flowing at 4 m/s. In 1 s you will move 3 m across the stream through the water and at the same time the stream will carry you down 4 m. Fig. 3.5 shows these velocities. AB represents the 3 m/s and BC the 4 m/s. The line AC is the resultant velocity and the actual path you will travel while in the boat. Notice the rule for adding velocities is the same as the rule for adding forces.

*Quantities which have direction as well as magnitude are called vectors.
Quantities which have only magnitude are called scalars*

fig 3.5 *the vector AB added to the vector BC is equivalent to the vector AC*



Velocities and forces are examples of vector quantities and must be added by the law of vector addition as follows:

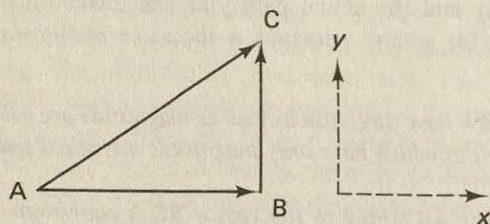
Represent the two vectors in magnitude and direction by the sides AB and BC taken in sequence round a triangle ABC. Then the resultant vector is represented in magnitude and direction by the third side AC

Mass and energy are examples of scalar quantities which add by simple arithmetic. A mass of 2 g added to a mass of 3 g always produces a total of 5 g. But forces of 2 N and 3 N do not always produce the same resultant; that depends on the direction in which each force acts.

3.3 RESOLVING FORCES

In the previous section we added two forces and found their resultant force. The opposite process can be carried out, namely, splitting up a single force into two component forces which produce exactly the same effect as the single force. The force AC in Fig. 3.6 may be resolved into

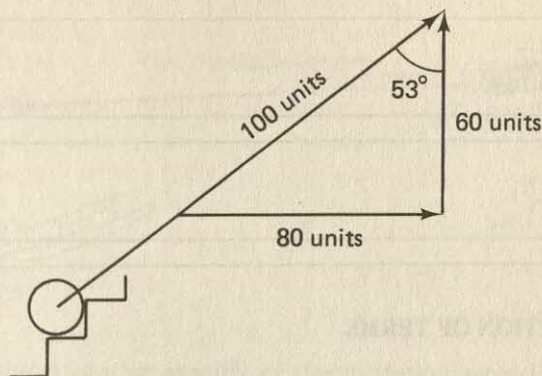
fig 3.6 the force AC may be resolved into two components AB and BC



two forces AB and BC at right-angles to each other. By this we mean that if two forces represented by AB and BC acted on a body they would have exactly the same effect as one force represented by AC. The force represented by AB is called the resolved component of AC in the direction x and the force represented by BC is called the resolved component of AC in the direction y.

As an example of this, imagine a man pulling a garden roller up some steps. Suppose he is pulling with a force of 100 N at an angle of 53° to the vertical. Fig. 3.7 shows how this force may be resolved into two

fig 3.7 the force of 100 units may be resolved into two components, a horizontal one of 80 units and a vertical one of 60 units



forces at right-angles. The components are 80 N and 60 N as indicated in the diagram. The effective force lifting the roller up the steps is therefore 60 N. He can make this effective force bigger by decreasing the angle made by the handle of the roller with the vertical.

MOTION

4.1 DEFINITION OF TERMS

Velocity is the distance travelled in unit time in a particular direction, or

$$\text{Velocity} = \frac{\text{Distance travelled}}{\text{Time taken}} \quad \text{in the specified direction} \quad (4.1)$$

A common unit is metres per second, written m/s, in a stated direction.

In everyday speech the terms velocity and speed are sometimes used to mean the same thing. However in science they mean something different. Speed is a scalar quantity and has no direction, whereas velocity is a vector quantity and should always have a direction stated.

Acceleration is the change in velocity in unit time, or

$$\text{Acceleration} = \frac{\text{Change in velocity}}{\text{Time taken for change}} \quad (4.2)$$

A common unit for acceleration is m/s^2 (in the direction in which the velocity changes).

4.2 NEWTON'S FIRST LAW OF MOTION

From everyday observation it might appear that we must exert a steady force to keep a body moving with constant velocity. For example, the water skier in Fig. 4.1 is being pulled by a constant force and moves with a constant velocity. The incompleteness of this argument is that we have neglected the force which the water exerts on the skis. The force exerted by the water on the skis and the force on the rope are equal but they act in opposite directions. Hence the total (or resultant) force on the skier is zero and he continues to move with constant velocity. Should the rope break, or the skier drop it, the force of the water on the skis quickly brings him to rest.

fig. 4.1 *constant speed seems to require a constant force acting on the skier*

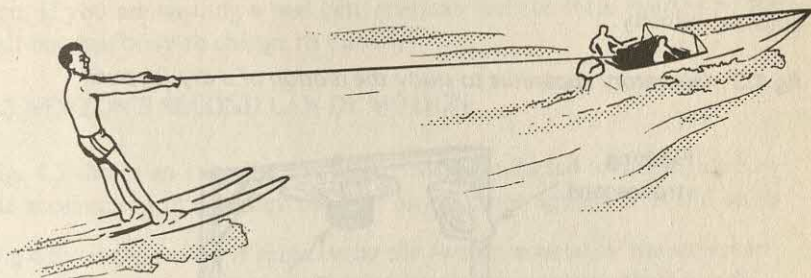
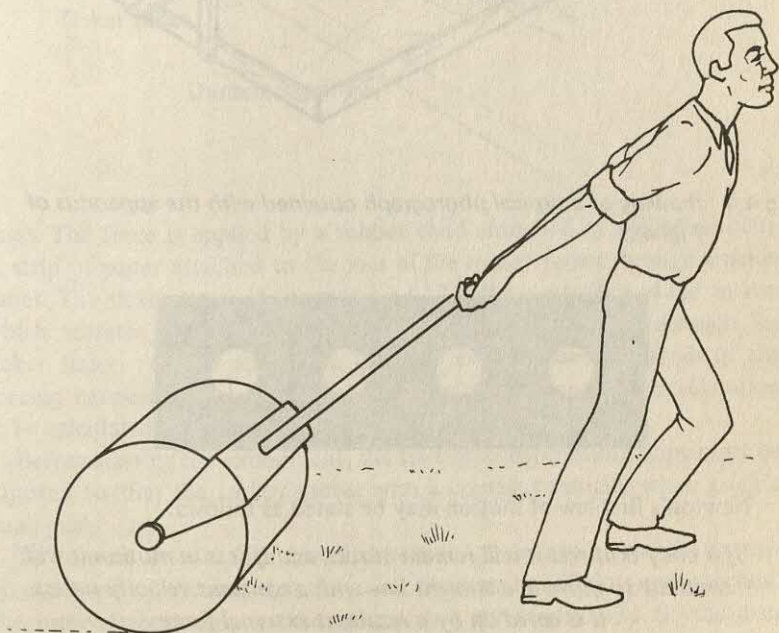


Fig. 4.2 shows another example of steady motion appearing to need a steady force. In this case it is the frictional forces which act in the opposite direction to the pull from the man, and again the resultant force is zero and the roller moves with a constant velocity.

One of the nearest approaches to motion on a frictionless surface in the laboratory is a dry-ice puck moving on a flat glass surface, where the

fig 4.2 *does a constant roller speed require a constant force?*



puck floats on a cushion of carbon dioxide gas. When the puck is given a push and photographed in the flashing light of a stroboscope (Fig. 4.3), the resulting picture (drawn in Fig. 4.4) shows that the puck moves a constant distance between each flash of light, that is, it is moving with a constant velocity.

fig 4.3 *laboratory apparatus to study the motion of a dry ice puck*

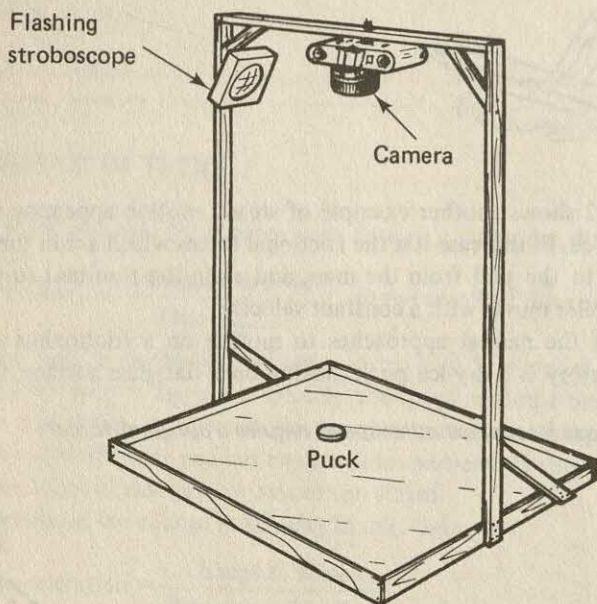
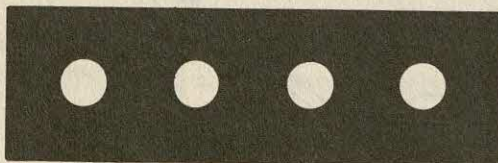


fig 4.4 *drawing of a typical photograph obtained with the apparatus of Fig. 4.3*



Newton's first law of motion may be stated as follows:

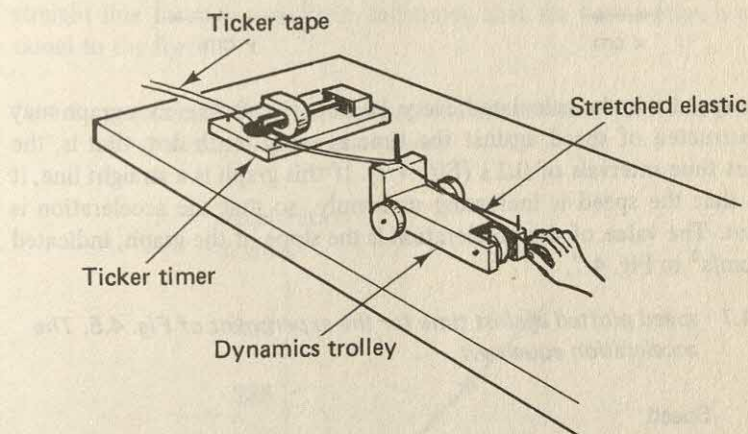
If a body is at rest it will remain at rest, and if it is in motion it will continue to move in a straight line with a constant velocity unless it is acted on by a resultant external force

We have all had experience of this law when travelling in a motor car. If you are in a car when the brakes are suddenly applied, you feel as if you are being thrown forwards. In fact, you are simply trying to move on with your constant velocity and it is the seat which is being pulled away from you. If you are wearing a seat belt, you can feel the force exerted by the belt on your body to change its velocity.

4.3 NEWTON'S SECOND LAW OF MOTION

Fig. 4.5 shows an experiment which is being conducted to investigate how the acceleration of a trolley depends on the force applied to it and on its

fig 4.5 *as the trolley is acted on by the force produced by the stretched elastic its motion down the friction-compensating slope is indicated by dots on the tape*



mass. The force is applied by a rubber band stretched by a fixed amount. A strip of paper attached to the rear of the trolley passes through a ticker timer. The ticker timer is operated by a 12 volt a.c. supply and has an arm which vibrates 50 times every second. As the tape moves through the ticker timer, 50 dots appear on the tape every second. Analysis of the spacing between the dots allows the velocity and hence the acceleration to be calculated.

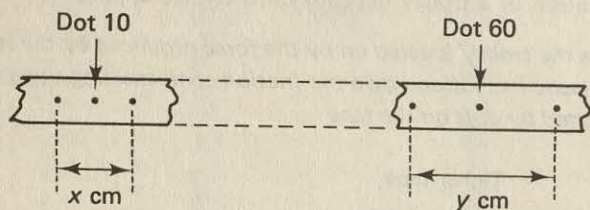
Before starting the experiment, the friction-compensating slope must be adjusted so that the trolley moves with a constant velocity when given a small push.

To convert the distance travelled between dots into a velocity in cm/s we make use of the fact that the ticker timer makes 50 dots every second. The intervals between two adjacent dots is therefore 0.02 s. So the time

to produce two such dot intervals, say between dot 9 and dot 11 if they are numbered consecutively on the tape, is $2 \times 0.02 = 0.04$ s.

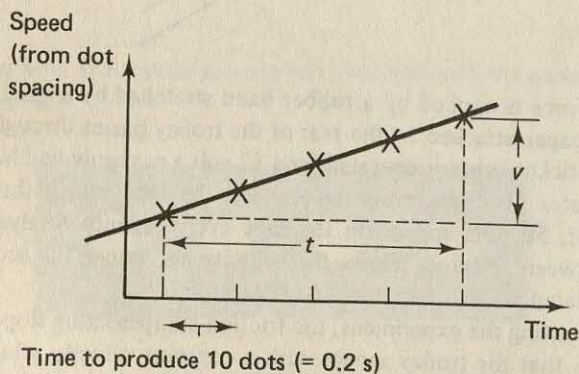
The acceleration may then be calculated as follows. Number the dots along the length of tape produced in the experiment. Suppose the distance between dots 9 and 11 is x cm (Fig. 4.6). Then the (average) speed at dot 10 is $x/0.04$ cm/s = $25x$ cm/s. Similarly if the distance between dots 59 and 61 is y cm, the speed at dot 60 is $y/0.04$ cm/s = $25y$ cm/s.

fig 4.6 the dots are made at 0.02 s intervals. The speed at dot 10 is $x/0.04 = 25x$ cm/s. The speed at dot 60 is $y/0.04 = 25y$ cm/s



The speed may be calculated every 10 dots (i.e. 0.2 s); so a graph may be constructed of speed against the time at every tenth dot, that is, the speed at time intervals of 0.2 s (Fig. 4.7). If this graph is a straight line, it means that the speed is increasing uniformly, so that the acceleration is constant. The value of the acceleration is the slope of the graph, indicated as v/t cm/s² in Fig. 4.7.

fig 4.7 speed plotted against time for the experiment of Fig. 4.5. The acceleration equals v/t



Alternatively, if it is assumed that the acceleration is uniform, its value may be calculated from the speeds at dots 10 and 60. The time interval between the production of these dots is exactly $50 \times 0.02 = 1.0$ s. So

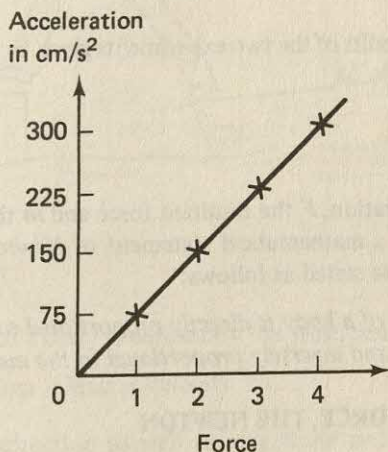
$$\begin{aligned}\text{Acceleration} &= \frac{\text{Change in speed}}{\text{Time taken for change}} \\ &\quad (1.0 \text{ s in this case}) \\ &= \frac{25y - 25x}{1.0} = 25(y - x) \text{ cm/s}^2\end{aligned}$$

(a) The relationship between force and acceleration

We call the force exerted by one elastic band one unit, the force from two elastic bands stretched the same amount two units, and so on. By applying forces of one, two, three and four units to the trolley in turn and analysing the tapes of the resulting motion, the acceleration may be calculated in each case, using one of the two methods described above.

The results of doing this for different forces are shown in Fig. 4.8, where the acceleration is plotted against the applied force. The graph is a straight line through the origin, indicating that the acceleration is proportional to the force.

fig 4.8 *graph indicating the relationship between acceleration and applied force acting on a constant mass*

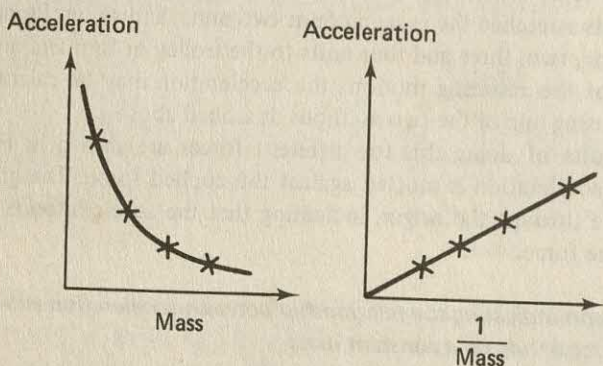


(b) The relationship between mass and acceleration

In this experiment the force is kept constant by using two rubber bands at constant stretch throughout the experiment. The mass may be changed by stacking one trolley on top of another. We say that one trolley has a mass of one unit, two trolleys stacked up a mass of two units, and so on. In each case the acceleration of the trolley stack along the slope is calculated

as in the previous experiment. Fig. 4.9 shows graphs of acceleration against mass and acceleration against $1/\text{mass}$. The graph of acceleration against $1/\text{mass}$ is a straight line through the origin, showing that the acceleration is inversely proportional to the mass.

fig 4.9 *graphs indicating the relationship between acceleration and mass for a constant applied force*



Combining the results of the two experiments given in the two preceding paragraphs we have

$$a \propto \frac{F}{m} \quad (4.3)$$

where a is the acceleration, F the resultant force and m the mass.

Equation (4.3) is a mathematical statement of *Newton's second law of motion*, which may be stated as follows:

The acceleration of a body is directly proportional to the resultant force acting on it and inversely proportional to the mass of the body

4.4 THE UNIT OF FORCE, THE NEWTON

One newton (1 N) is the force which gives a mass of one kilogram (1 kg) an acceleration of one metre per second per second (1 m/s²)

Defining the newton in this way enables us to write equation (4.3) as

$$a = \frac{F}{m} \quad \text{or} \quad F = ma \quad (4.4)$$

where F is in newtons, m in kilograms and a in metres/second².

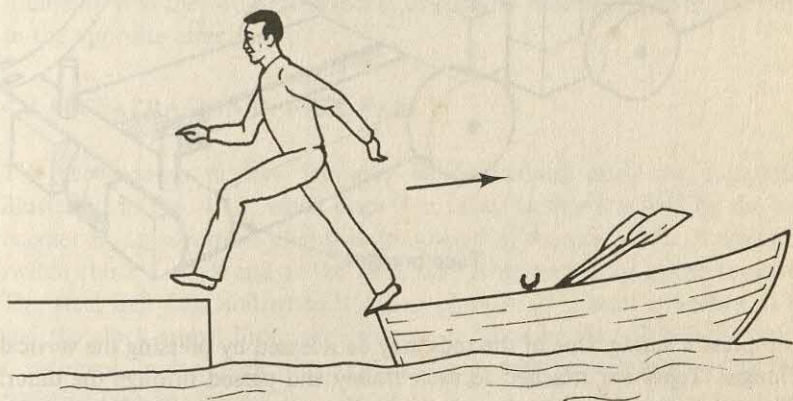
4.5 NEWTON'S THIRD LAW OF MOTION

Newton's third law states:

If a body A exerts a force on a body B then B exerts an equal and opposite force on body A

An appreciation of the third law will save you from getting very wet when stepping out of a rowing boat (Fig. 4.10)! As you step out of the boat, the force which the boat exerts on you (enabling you to move forwards) means that there is an equal force acting backwards on the boat causing it to accelerate backwards, so that the distance between the boat and the pier increases rapidly.

fig 4.10 a demonstration of Newton's third law of motion



4.6 MOMENTUM

The momentum of a body is defined by the relationship

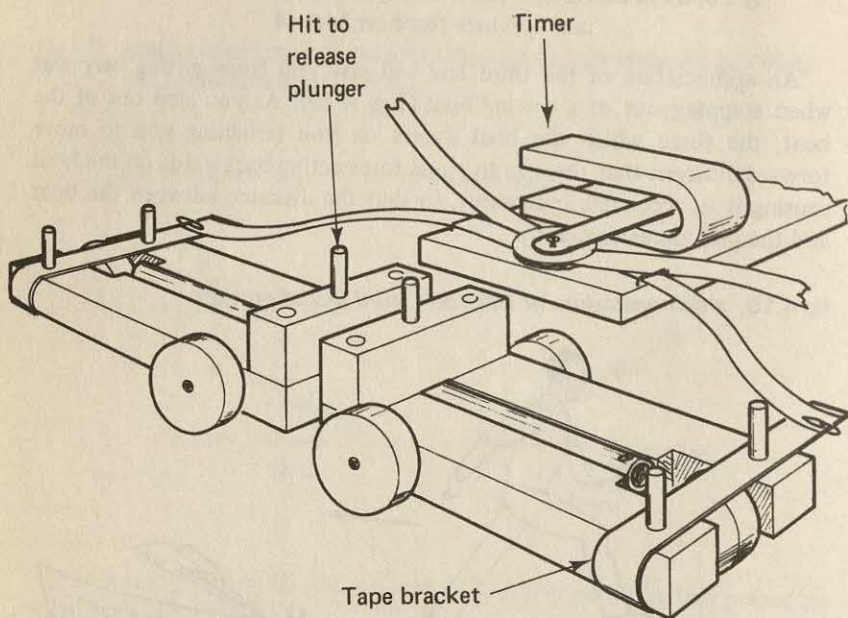
$$\text{Momentum} = \text{Mass} \times \text{Velocity} \quad (4.5)$$

Momentum has direction as well as magnitude and is therefore a vector quantity (see also section 3.2). For a group of bodies the total momentum is equal to the sum of the individual momenta.

4.7 CONSERVATION OF MOMENTUM

An important result may be obtained using the apparatus illustrated in Fig. 4.11. The trolleys have protruding rods which may be pushed in to

fig 4.11 *the arrangement of two trolleys and ticker timer for the so-called 'explosion' experiment. The masses of the trolleys may be varied by adding loads to them. (Redrawn from material kindly supplied by Philip Harris Ltd)*



compress a spring. One of the rods may be released by pressing the vertical plunger. Tapes are attached to each trolley and passed through the timer. When the plunger is struck and the trolleys move apart, their velocities may be analysed by analysis of the tapes. The results show that if one trolley with mass m_1 moves off with velocity v_1 in one direction, the other trolley of mass m_2 moves off with velocity v_2 in the other direction, where

$$m_1 v_1 = m_2 v_2$$

that is, the momentum gained by one trolley as a result of the explosion is equal and opposite to the momentum gained by the other.

In another kind of experiment a trolley may be made to run into the back of another initially stationary one, the two being made to stick together using a pin on the first which sticks into a lump of plasticine on the second. In each case, no matter what the masses of each trolley, it is found that the total momentum before impact (i.e. the momentum of the moving trolley) is equal to the total momentum after impact (i.e. the momentum of both trolleys added together).

These experiments are examples of the *law of conservation of momentum* which states that

If no external force acts on a system in a particular direction then the total momentum of the system in that direction remains unchanged

4.8 ROCKET AND JET MOTORS

In both rocket and jet motors a fuel is burnt in a combustion chamber. The difference between them is that in a rocket the gas in which the fuel is burnt is carried (often in the form of liquid oxygen) in the rocket itself; in a jet engine the fuel is burnt in air drawn into the engine from an intake at the front end of the engine.

As a result of combustion, the gases reach a high temperature and pressure. They are expelled from the exhaust nozzle at high speed, gaining momentum as they do so. The rocket or jet must therefore gain momentum in the opposite direction.

4.9 ACCELERATION IN FREE FALL

The acceleration in free fall may be determined using the apparatus illustrated in Fig. 4.12. When plate A is raised so that it is held by the bar magnet B, an electrical circuit is completed at the contact C. A separate switch starts a clock and at the same time switches off the electromagnet. The steel ball falls and when it strikes plate A the circuit is broken at C and the clock stops. From the distance, s , fallen by the ball and the time, t , recorded on the clock, the acceleration, g , may be calculated using the equation (see Section 4.12)

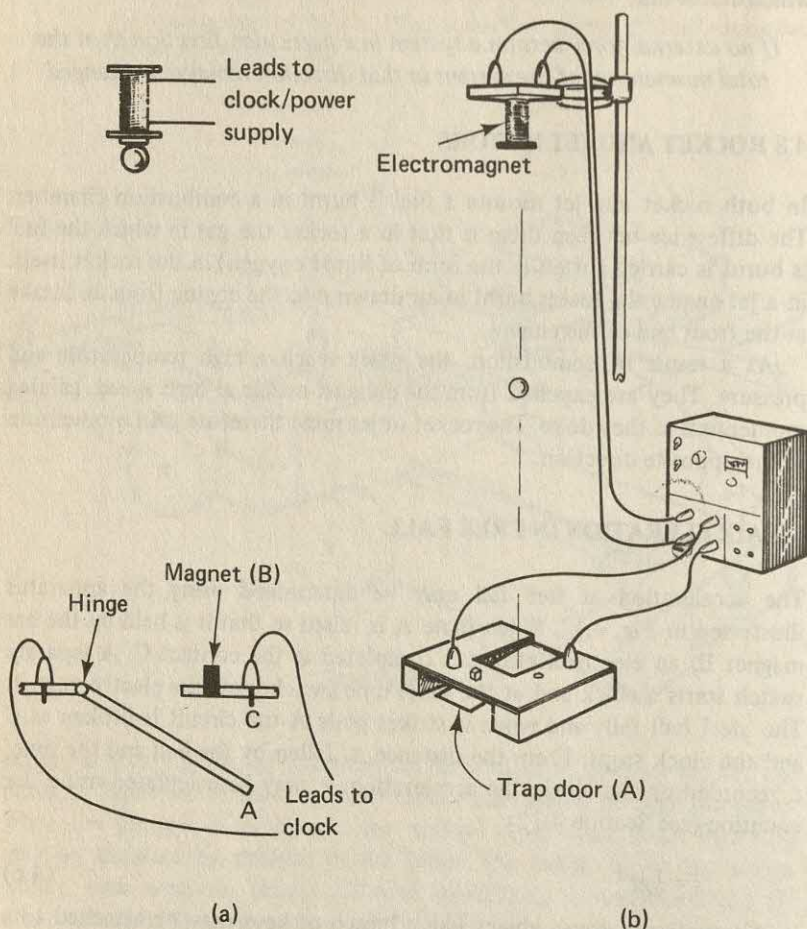
$$s = \frac{1}{2}gt^2 \quad (4.6)$$

Alternatively, some object like a bunch of keys may be attached to a ticker tape and the ticker timer fixed so that the tape can fall vertically through it (Fig. 4.13). The mass is released and subsequent analysis of the tape enables the acceleration due to gravity to be calculated.

The acceleration due to gravity, in free fall, is about 10 m/s^2 (or more exactly 9.8 m/s^2) on Earth. It has a quite different value on the surface of the Moon, or elsewhere in the Solar System.

4.10 FRICTIONAL FORCES IN SOLIDS

Whenever one surface slides over another surface, there is always a force opposing motion. The force results from the fact that no surfaces are perfectly smooth. However highly polished the surfaces, on a molecular

fig 4.12 the magnet/clock apparatus for measuring g 

scale they might look something like Fig. 4.14. If an attempt is made to slide one surface over the other, the tiny projections on each surface will catch on each other and this results in a force opposing motion. This force is called *friction*.

One basic property of frictional forces may be investigated using the apparatus illustrated in Fig. 4.15. A wooden block rests on a horizontal bench top and has a hook on one side to which a spring balance is attached. When the spring balance is pulled sideways, a horizontal force will be exerted on the block, trying to make it move off across the bench top. At first nothing happens, although the spring balance force steadily increases. Eventually the block will start to move and the spring balance reading just before this happens indicates what is called the *limiting static friction*.

fig 4.13

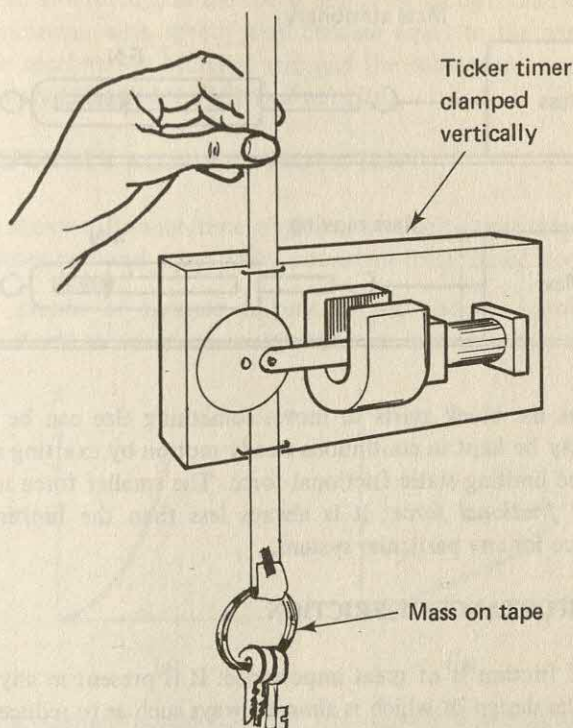
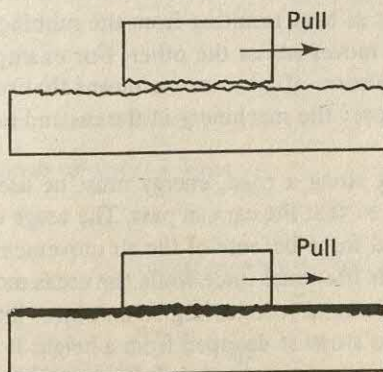
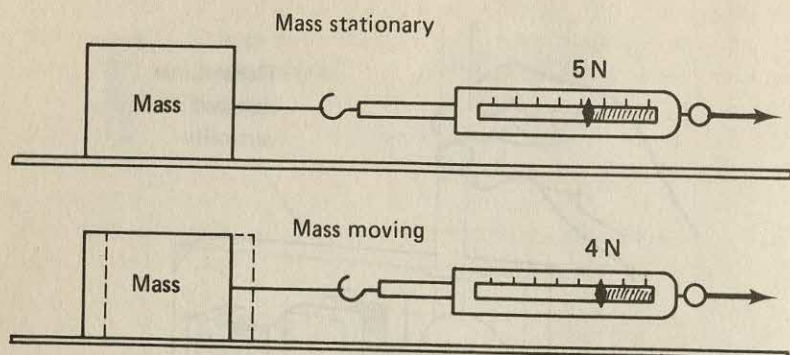
fig 4.14 *an oil film separates the two surfaces and reduces the frictional force*

fig 4.15 an apparatus with which frictional forces can be investigated



As soon as the block starts to move, something else can be noticed. The block may be kept in continuous steady motion by exerting a smaller force than the limiting static frictional force. The smaller force is termed the *dynamic frictional force*; it is always less than the limiting static frictional force for any particular system.

4.11 THE IMPORTANCE OF FRICTION

The force of friction is of great importance. It is present in any moving machinery, the design of which is almost always such as to reduce friction as much as possible.

There are two main ways of reducing friction, either by separating the two parts in relative motion using perhaps a film of oil or grease, or else by introducing something which will roll between the surfaces; ball bearings or roller bearings have this function.

In machines it is important to reduce friction as much as possible, otherwise energy is lost as heat resulting from the rubbing together of the surfaces, one of which moves across the other. For example, any friction in the engine or transmission of a motor car means that some petrol must be used up in turning over the machinery in the car and not in driving the car along the road.

While a car is going along a road, energy must be used up in moving the air out of the way so that the car can pass. The usage of energy means that there is a frictional force because of the air movement necessary, and work is done against this frictional force while the car is moving.

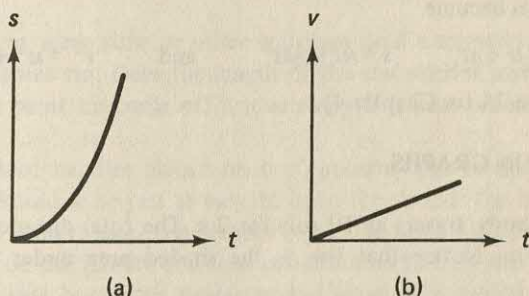
There is a similar frictional force acting on an object like a stone falling through the air. When a stone is dropped from a height it initially accelerates at about 9.8 m/s^2 , as previously stated, because the only force on it

initially is its weight. Subsequently, as its speed increases there is a frictional force, due to the air resistance, acting in the opposite direction to the weight. Eventually, as the speed builds up enough, the frictional force (which increases with speed) may become equal to the weight. At this point the acceleration becomes zero and the subsequent, constant downward velocity is termed the *terminal velocity*.

4.12 UNIFORMLY ACCELERATED MOTION

Fig. 4.16 shows a distance/time graph and a velocity/time graph for a body *starting from rest* and acted on by a constant force. Since a constant force

fig. 4.16 *graphs of distance (s)/time (t) and velocity (v)/time (t) for a body acted on by a constant force*



produces a constant acceleration (a), the velocity/time graph is a straight line of gradient a . The distance/time graph is a curve because the distance travelled is proportional to t^2 . This may be shown as follows:

From equation (4.2) we have

$$a = \frac{v}{t}$$

or

$$v = at \quad (4.7)$$

and the distance travelled s is given by

$$s = (\text{Average velocity}) \times \text{Time}$$

But

$$\begin{aligned} \text{Average velocity} &= \frac{\text{Initial velocity} + \text{Final velocity}}{2} \\ &= \frac{0 + at}{2} = \frac{1}{2}at \end{aligned}$$

Therefore

$$s = \frac{1}{2}at \times t = \frac{1}{2}at^2$$

i.e.

$$s = \frac{1}{2}at^2 \quad (4.8)$$

Hence s is proportional to t^2 and the graph of s against t is a curve as shown in Fig. 4.16.

Eliminating t from equations (4.7) and (4.8) we get

$$v^2 = 2as \quad (4.9)$$

Problems about bodies moving with a constant acceleration can often be quickly solved by using the above equations.

If the body does not start from rest but has an initial velocity u , then the equations become

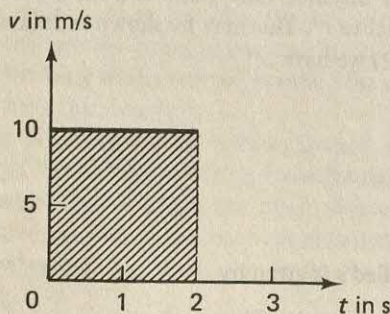
$$v = u + at, \quad s = ut + \frac{1}{2}at^2 \quad \text{and} \quad v^2 = u^2 + 2as$$

(see question 14 for Chapter 4).

4.13 MOTION GRAPHS

Suppose a body travels at 10 m/s for 2 s. The total distance travelled is $10 \times 2 = 20$ m. Notice that this is the shaded area under the graph in Fig. 4.17.

fig 4.17 the shaded area is the distance travelled



In general we may say that

For a graph of velocity/time, (i) the gradient is the acceleration, and
(ii) the area under the graph is the total distance travelled

For a graph of distance/time, the gradient of the graph at any point
is the velocity of the body at that point

LEVERS AND CENTRES OF GRAVITY

5.1 LEVERS IN EVERYDAY LIFE

Most of you at some time or other will have used a screwdriver to prize the lid off a paint tin. Does the length of the screwdriver matter? Would the lid of the paint tin come off just as easily if it were prised off using a coin?

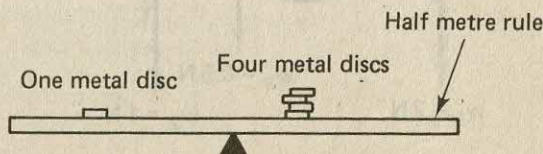
Why are door handles placed on the opposite edge of the door from the hinges? Would it be just as easy to open the door if the handle were nearer to the hinge? Suppose someone tries to close an open door by pushing hard on the handle and someone else tries to stop him by pushing on the other side but much nearer to the hinge. Who would win? What other factor besides the force determines the turning effect of the force?

This chapter deals with the principles which determine the answers to these questions.

5.2 AN EXPERIMENT TO INVESTIGATE THE TURNING EFFECT OF A FORCE

A half metre rule (or any other beam) is balanced at its centre. Place four metal discs (UK 10p pieces are suitable) so that their centres are 5 cm from the *fulcrum* (i.e. the point of pivot). Now see where one disc must be placed on the other side so that the beam is in *equilibrium* (i.e. the beam is balanced) (Fig. 5.1).

fig 5.1 *an experiment to investigate the law governing the balancing of a beam*



Next balance the four discs using two discs on the other side, then three and finally four discs. The results of such an experiment are shown in Table 5.1.

Table 5.1 *results for the experiment illustrated in Fig. 5.1*

<i>No. of discs (units of force) on left-hand side</i>	<i>Distance from fulcrum (cm)</i>
1	20.0
2	10.0
3	6.7
4	5.0

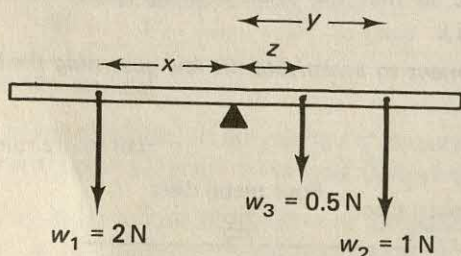
The four discs on the right-hand side try to turn the beam in a clockwise direction. The discs on the left-hand side tend to turn the beam anticlockwise. Clearly the turning effect of the force depends on its distance from the fulcrum. But how can we actually work out the turning effect? The results in Table 5.1 show that the number of discs (units of force) multiplied by the distance from the fulcrum is a constant, in this case equal to 20.

We can therefore measure the turning effect of a force by multiplying together the force and the perpendicular distance from the fulcrum to the line of action of the force. Thus

$$\text{Moment of a force} = \text{Force} \times \text{Perpendicular distance from the fulcrum to the line of action of the force}$$

The experiment illustrated in Fig. 5.2 is designed to investigate further the turning effects of forces. A number of weights are hung from a metre rule and their distances from the fulcrum adjusted until the rule is horizon-

fig 5.2 *a more complex experiment, with two forces on one side of the fulcrum balancing one on the other side*



tal. The results of such an experiment are shown in Table 5.2. In Table 5.3 the anticlockwise moment for this experiment, $W_1 \times x$, and the two clockwise moments, $W_2 \times y$ and $W_3 \times z$, are calculated. In the readings shown the anticlockwise moment is kept constant and it can be seen that the total clockwise moment is obtained by adding the two separate moments.

Table 5.2 *results for the experiment illustrated in Fig. 5.2*

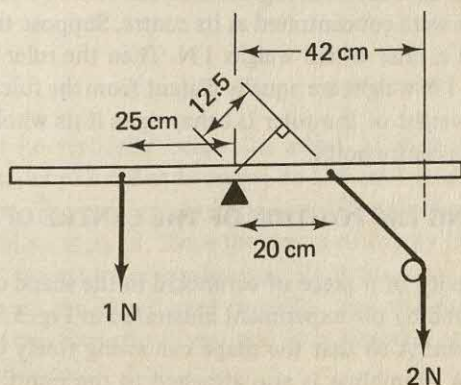
W_1 (N)	W_2 (N)	W_3 (N)	x (cm)	y (cm)	z (cm)
2	1	0.5	20	35	10
2	1	0.5	20	25	30
2	1	0.5	20	20	40

Table 5.3 *calculations for the results of the experiment (Fig. 5.2) given in Table 5.2*

$W_1 x$ (anticlockwise moment)	$W_2 y + W_3 z$ (total clockwise moment)
40	$35 + 5 = 40$
40	$25 + 15 = 40$
40	$20 + 20 = 40$

Fig. 5.3 shows an arrangement where the forces acting on the ruler are not parallel to one another. A pulley is used to change the direction of the 2 N force. The position of the pulley is adjusted until the ruler is horizontal. The readings obtained in a particular experiment are shown

fig 5.3 *balancing a beam when the forces acting on it are not parallel*



on the diagram. The anticlockwise moment is $1 \times 25 = 25 \text{ N cm}$. Which distance must we multiply the 2 N force by, in order to get 25 N cm ? You can see from the diagram that the required distance, which is 12.5 cm , is the perpendicular distance from the fulcrum to the *line of action* of the force.

We can summarise the results of the above experiments by the following statements:

The moment of a force about a point is measured by the product of the force, and the perpendicular distance from the point to the line of action of the force

When a body is in equilibrium, the sum of the clockwise moments acting on the body about a point is equal to the sum of the anticlockwise moments about the same point

5.3 CENTRE OF GRAVITY

A ruler can be balanced if it rests on a support under its centre (Fig. 5.4). Every part of the ruler is attracted by the Earth, but the ruler behaves as

fig 5.4 a uniform ruler balanced at its midpoint



if its whole weight were concentrated at its centre. Fig. 5.5 illustrates an experiment you can conduct with a friend. A thread with a loop on the end is tied to the centre of a ruler and a weight of equal mass also has a string with a loop attached to it. Provided your friend keeps his eyes shut, he will not be able to tell whether it is the metre rule or the weight which you are putting on to his finger.

The experiment illustrated in Fig. 5.6 also shows that a ruler behaves as if its whole weight were concentrated at its centre. Suppose the experiment is conducted with a ruler which weighs 1 N . Then the ruler will balance if its centre and the 1 N weight are equally distant from the fulcrum, as shown in Fig. 5.6. The weight of the ruler is behaving as if its whole weight were concentrated at its centre point.

5.4 DETERMINING THE POSITION OF THE CENTRE OF GRAVITY

The centre of gravity of a piece of cardboard in the shape of England and Wales may be found by the experiment illustrated in Fig. 5.7. A small hole is made at the point A so that the shape can swing freely on a horizontal pin through A. A plumbline is also attached to the point. The centre of

fig 5.5 *the ruler behaves as if its whole weight were concentrated at its centre of gravity*

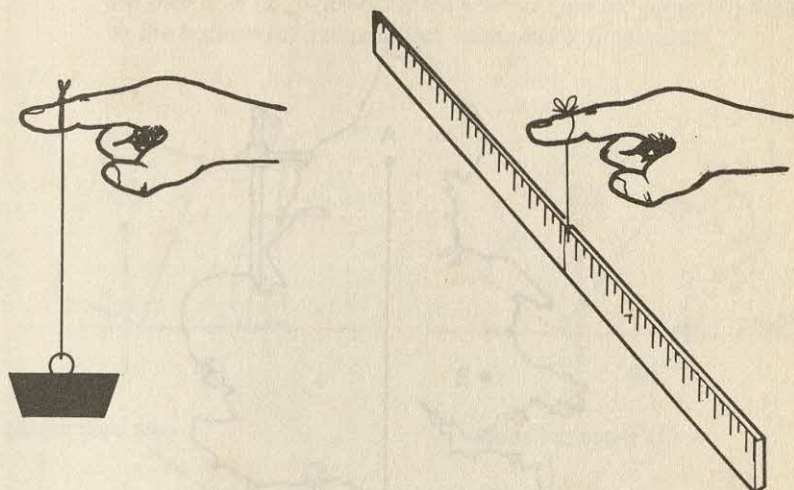
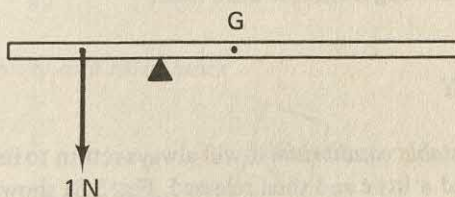
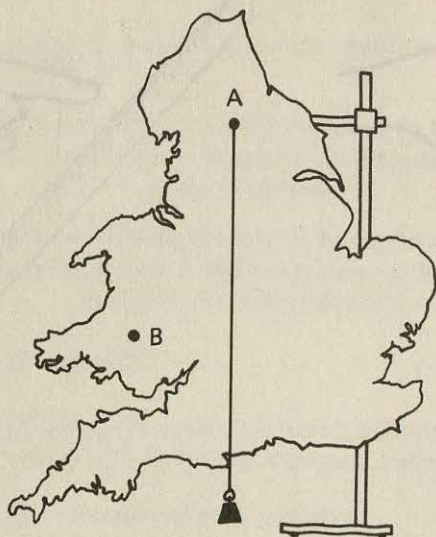


fig 5.6 *balancing a ruler off-centre using just one applied force*



gravity must lie vertically below the point A, when the card is hanging. (Can you see why?) A line is drawn on the card to show the position of the plumbline. The card is then hung from the hole B and the new position of the plumbline marked. Since the centre of gravity lies on the plumbline in each case, the centre of gravity must lie at the point where the two lines cross. If you do the experiment carefully, you will find that the card will balance on your fingertip if you put your fingertip at the point of intersection of the lines.

fig 5.7 a freely suspended object rotates until its centre of gravity is vertically below its point of suspension



Do you think the accuracy of the experiment would be improved if a third hole were made and the direction of the plumbline marked on the card when it was hung from this third hole?

5.5 STABILITY

If a body is in *stable equilibrium* it will always return to its original position if it is displaced a little and then released. Fig. 5.8a shows a body in stable equilibrium since its weight gives rise to a moment which will cause it to return to its original standing position when released. In Fig. 5.8b the body is *unstable* since when it is released its weight gives rise to a moment which makes it topple over. Two ways of making a body more stable are illustrated in (c) and (d). A body is said to be in *neutral equilibrium* if when displaced it stays in its new position, i.e. a ball on a horizontal table.

Consider the experiment illustrated in Fig. 5.9. A wooden block has a plumbline attached to a point on a horizontal line through its centre of gravity. The block is standing on a rough surface which can be tilted. The block will not topple over provided the plumbline crosses the base between A and B.

fig 5.8 *stable and unstable equilibrium. The object in (a) is stable because it will return to its original position when released. In (b) when released it will topple further and is said to be unstable. Enlarging the base area (c), or lowering the centre of gravity by adding ballast to the bottom (d) makes a body more likely to be stable*

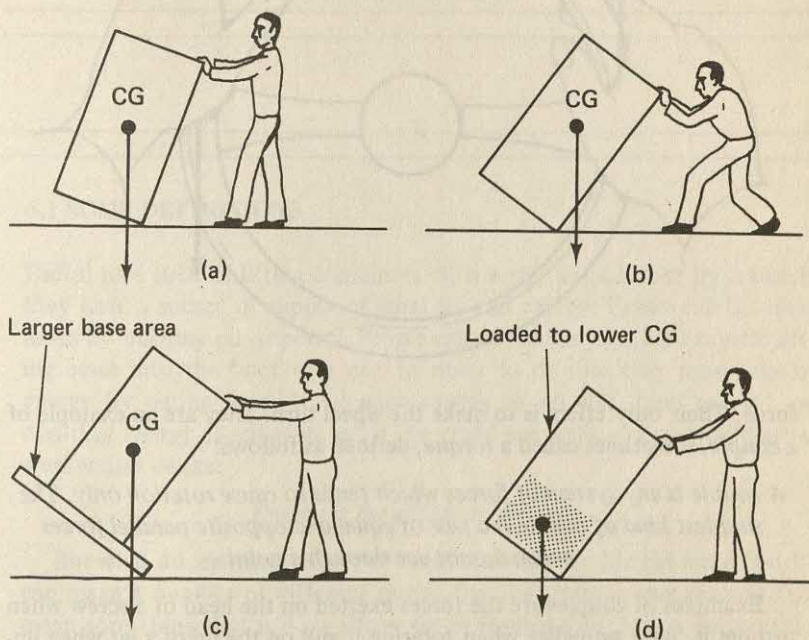
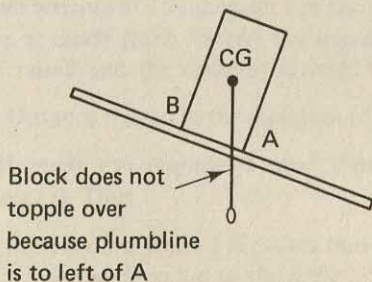


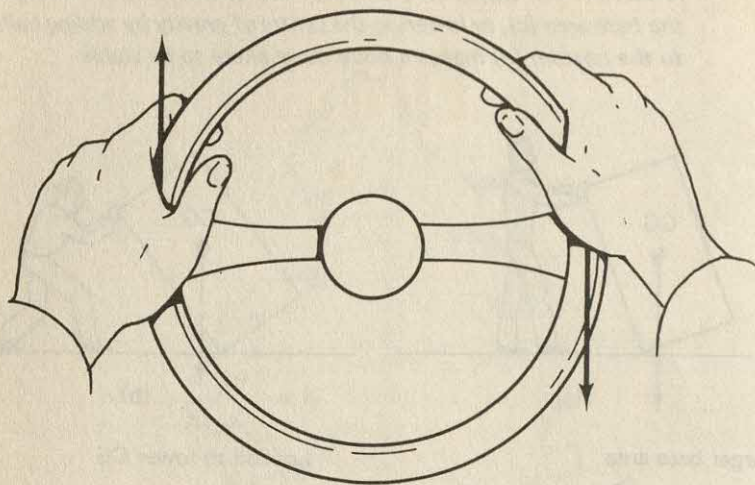
fig 5.9 *the stability of a tilted block*



5.6 COUPLES

Suppose you grip a steering wheel with your two hands and turn it. The two forces are as indicated in Fig. 5.10. They cannot be replaced by a single

fig 5.10 *forces exerted by the hands make the steering wheel rotate. These forces together make a couple*



force. Their only effect is to make the wheel turn. They are an example of a *couple*, sometimes called a *torque*, defined as follows:

A couple is any system of forces which tends to cause rotation only. The simplest kind of couple is a pair of equal and opposite parallel forces which do not act through a point

Examples of couples are the forces exerted on the head of a screw when turning it, on a propeller when rotating it and on the lid of a jar when unscrewing it.

WORK, ENERGY AND POWER

6.1 SOME DEFINITIONS

Useful jobs such as lifting containers onto a ship can be done by cranes if they have a source of supply of what we call energy. Cranes can lift their loads by burning oil or petrol. People can do similar jobs, for example lifting cases into the boot of a car. In order to do this, they must take in energy by eating food. If we were unable to get this 'food energy', or *chemical energy* as scientists call it, we could not work and we would grow weaker and weaker.

Energy is the ability to do work

But what do scientists mean by 'work'? In everyday life the word 'work' can mean a number of different things, but the scientist uses the word to mean something that is done whenever an applied force causes movement. Work is done whenever a force moves its point of application. For example, if you lift a load of 1 N onto a shelf 1 m higher, you do a certain amount of work. If you lift two weights of 1 N through 1 m or a weight of 2 N through 1 m, you do twice as much work. Clearly the quantity of work depends on both the load raised and the distance moved. We define 'work' as:

$$\text{Work} = \text{Force} \times \text{Distance moved in the direction of the force} \quad (6.1)$$

To move 1 N through 1 m requires 1 newton-metre (N m) of work. 1 N m is called a joule (J). Thus

*1 J of work is done when 1 N moves through 1 m
in the direction of the force*

6.2 VARIOUS FORMS OF ENERGY

We have already stated that oil and petrol can be used to supply energy to a crane to enable it to lift a load. Other forms of energy could have been

used to work the crane. For example, electrical energy could drive an electric motor or heat energy could be used to drive a steam engine. In one of the cases, electrical energy is being used to do work and, in the other case, heat energy is being used to do work.

There are many different kinds of energy. Many machines have been devised which will convert one form of energy into another.

6.3 ELECTRICAL ENERGY

Batteries and dynamos produce *electrical energy*. This energy, possessed by any electric current flowing in a circuit, is a very clean and convenient form of energy. In a battery it is produced by chemical reactions (i.e. chemical energy is converted into electrical energy). In a generating station the initial source of energy which the dynamo converts into electrical energy may be chemical energy (from coal or oil being burnt) or *nuclear energy* (from uranium or other nuclear fuel).

Electrical energy may be converted into *heat energy* in an electric fire, or into *light energy* (together with some heat as well) in a bulb.

6.4 KINETIC ENERGY (KE)

Kinetic energy is the energy possessed by a moving body

Suppose a stone drops over the side of a cliff. As it falls, its kinetic energy increases as it goes faster and faster. It has maximum kinetic energy just before it hits the ground. When it hits the ground this energy is turned into *heat energy* and some *sound energy*. The kinetic energy of a moving car can demolish a brick wall. Hydroelectric power stations make use of the kinetic energy of water running down mountain sides. The kinetic energy of the water is used to drive a dynamo and thus produce electrical energy.

It may be shown mathematically that a body of mass m moving with speed v has a kinetic energy given by the expression $\frac{1}{2}mv^2$. Consider the body of mass m initially at rest and acted on by a force F which causes it to undergo an acceleration a . Suppose that it travels a distance s under this acceleration in time t .

Then the work done by the force on the body is given by

$$\text{Work (W)} = \text{Force} \times \text{Distance} = Fs$$

(from equation (6.1)). But

$$F = ma$$

(from equation (4.4)) and

$$s = \frac{1}{2}at^2$$

(from equation (4.8)), so that

$$W = (ma) \times (\frac{1}{2}at^2) = \frac{1}{2}m(at)^2$$

As the final velocity v is given by

$$v = at$$

(see equation (4.7))

$$W = \frac{1}{2}mv^2$$

The work done on the body is equal to the energy stored in the body in this case, so that the kinetic energy of this body is given by the expression

$$\text{Kinetic energy} = \frac{1}{2}mv^2 \quad (6.2)$$

The units of kinetic energy are the same as the units of work. If m is in kg and v is in m/s, then $\frac{1}{2}mv^2$ is in joules (J).

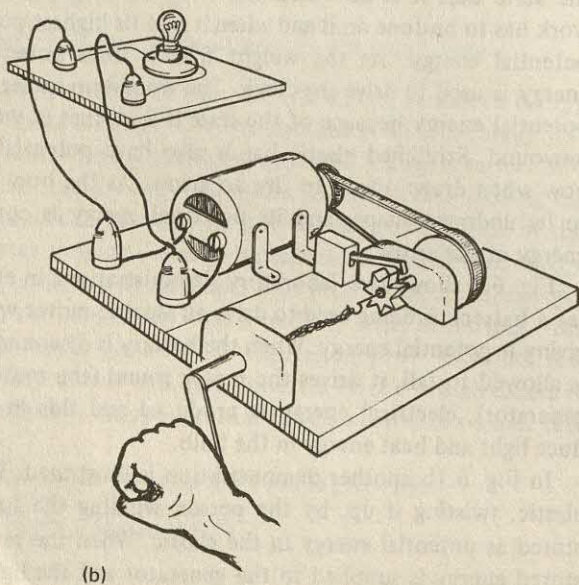
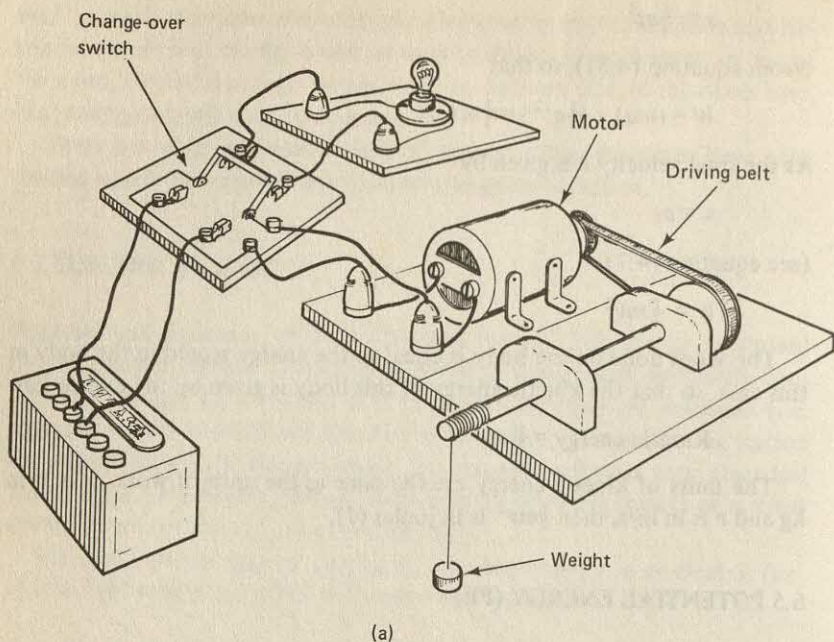
6.5 POTENTIAL ENERGY (PE)

Potential energy is the energy that a body has by virtue of its position or the state that it is in. When the weight in a grandfather clock is raised, work has to be done on it and when it is at its highest point it has maximum potential energy. As the weight falls it loses potential energy and the energy is used to drive the clock. The wound-up spring in a wristwatch has potential energy because of the state it is in, that is, wound up rather than unwound. Stretched elastic bands also have potential energy. So does a bow when drawn ready to fire an arrow. As the bow is released it returns to its undrawn shape, and its potential energy is converted into kinetic energy of the arrow.

Fig. 6.1 shows two laboratory demonstrations in energy conversion. In (a) a battery is being used to drive an electric motor which lifts the weight giving it potential energy. When the battery is disconnected and the weight is allowed to fall, it drives the motor round (the motor then behaves as a generator), electrical energy is produced and this in turn is used to produce light and heat energy in the bulb.

In Fig. 6.1b another demonstration is illustrated. Work is done on the elastic, twisting it up, by the person winding the handle. This energy is stored as potential energy in the elastic. When the ratchet is released this stored energy is supplied to the generator and the light bulb momentarily flashes.

The change in gravitational potential energy associated with a change in vertical height may be readily calculated. Suppose a stone has a mass m kg,



then it has a weight of mg N, where g is the force of gravity on each kilogram (about 10 N/kg). When the stone is raised to a height h m above

fig 6.1 two laboratory demonstrations of the conversion of energy. In (a) energy from a battery drives a motor which lifts a weight, giving it potential energy. This potential energy can then turn the motor (as a generator) producing electrical energy to light a bulb. In (b) work is done on the elastic and stored as potential energy. The stored energy can again be converted into electrical energy to light the bulb

the ground, the work done on it is given by

$$\text{Force} \times \text{Distance} = mg \text{ (N)} \times h \text{ (m)} = mgh \text{ (J)} \quad (6.3)$$

and this is the gain in potential energy when the stone is raised a height h m. If the stone falls again, this potential energy is converted into kinetic energy.

The velocity of the stone when it reaches the ground again may be calculated, for the loss in potential energy during the fall is equal to the gain in kinetic energy. Using the usual symbols and equation (6.2) for the kinetic energy, we therefore have

$$mgh = \frac{1}{2}mv^2$$

so that

$$v^2 = 2gh \quad \text{or} \quad v = \sqrt{2gh}$$

6.6 CONSERVATION OF ENERGY

In all our discussion in this chapter or elsewhere we have assumed that, when any work is done, the energy needed to do it has come from somewhere and that, after the work is done, the energy has been transformed into some other form.

All experiments have been found to be consistent with the law of conservation of energy which states that:

Energy cannot be created or destroyed; it can only be converted from one form to another

6.7 HEAT ENERGY

If energy cannot be created or destroyed, and if more and more energy is released on Earth by the burning of coal and oil, what becomes of this energy? The answer is that it mostly becomes heat energy. For example, in a motor car some of the chemical energy in the fuel becomes heat energy in the exhaust gases. Only a small amount (less than 30%) becomes mechanical energy and a lot of this becomes heat energy in the bearings. When

the car is stopped by applying the brakes, most of the kinetic energy of the car becomes heat energy in the wheels and brake shoes and this eventually gets lost to the atmosphere. Even while the car is going along at constant speed, energy is used up by the engine in driving the car through the air, and this becomes increased kinetic energy of the air molecules past which the car has gone (increased heat energy).

If this were the whole story, the air round the Earth would be gradually heating up as time passed. This is not happening because energy is being continually lost from the Earth's surface into space. The surface temperature of the Earth is determined almost solely by the energy radiated to the Earth from the Sun. The Earth's average surface temperature just keeps the total energy received from the Sun and the total energy radiated from Earth into space in balance. The energy released on Earth by the burning of various fuels, etc., makes very little difference to this balance.

6.8 POWER

What usually matters for any machine is not just how much work it can do, but the rate at which it can do the work.

The power of a machine is the rate at which it does work

From this definition it follows that power is measured in joules per second or watts, for

$$1 \text{ watt (W)} = 1 \text{ joule per second (J/s)}$$

Like any other metric unit, the watt has multiples and submultiples. One kilowatt (kW) equals 1000 watts. One megawatt (MW) equals 1000 000 watts.

When a 100 W light bulb is switched on it uses energy at a rate of 100 J/s. A power station might be rated at a maximum output of 500 MW. This means that it could light up five million 100 W light bulbs simultaneously.

It is possible to calculate the power of the human body when running up stairs (see question 8 for Chapter 6).

6.9 THE ENERGY CRISIS

Since the early 1970s there has been increasing talk of an energy crisis. It has become generally realised that the Earth's known reserves of oil are rapidly running out. If there had continued to be a 5% annual growth in world demand for oil, as there was in the 1960s, and even if technological advance had made possible a 100% recovery of oil from oil fields (rather

than about 50% as at present) then all known oil reserves would have been used up by about 2050.

A world recession reduces demand and makes the oil last longer, but there is still a need to reduce oil consumption as much as possible and to search for alternative forms of energy where possible.

(a) Energy conservation

Various measures being taken include better insulation of buildings, so that energy used for space heating in winter is not quickly lost to the outside. Car body shapes have been improved, so that less energy is wasted in driving the car through the air and the petrol consumption is improved. Where possible there is an attempt to move away from the use of oil to the use of coal or some other longer-lasting fuel (hence the argument for the increased electrification of the railways).

(b) Alternative forms of energy

The realisation that there is an energy crisis has spurred on the search for alternative sources of energy. The principal sources used at present are oil, coal, nuclear (fission), hydroelectric power and natural gas. Others which could become more widely used are as follows:

(i) Solar power

Special panels act like a miniature greenhouse and can be used, for example, to pre-heat the domestic hot water supply. Solar cells are used on spacecraft to convert heat and light energy from the Sun into electrical energy.

(ii) Wave power

Specially shaped devices have been designed which move as waves in the sea pass them. The movement can be used to drive dynamos.

(iii) Tidal power

The enormous movement of water round the world's coasts between high and low tides provides a very large source of energy. This has been tapped at La Rance, France, where there is a tidal barrage with a peak output of 400 MW. The difference in depth of water on either side of the barrage is used to drive a set of turbines. A proposed barrage across the Severn Estuary would be some 40 times larger than this, producing about 10% of the UK electrical energy demand, but at high capital cost.

(iv) Geothermal power

A few places in the world have the temperature near the surface sufficiently high to produce steam which is then used to drive turbines and generate electricity in the usual manner (e.g. The Geysers, California, USA).

(v) Wind power

The traditional windmill is well known and now uncommon. Newer types of windmill are being developed which could be a significant source of energy in exposed, windy places.

(vi) Fusion power

All existing nuclear power stations operate on the fission principle, the source of energy being the fission of appropriate nuclei (e.g. uranium, plutonium). Another type of nuclear reaction is the fusing together of hydrogen nuclei to form helium nuclei, with an associated large release of energy. This reaction takes place at an uncontrolled rate in a hydrogen bomb explosion; scientists are trying to control the reaction in order to produce a fusion power station. If they are successful, we would have a virtually unlimited source of energy because the world's oceans contain enough of the right type of hydrogen to last for many thousand million years.

WORKED EXAMPLE

The brakes of a new experimental very fast train give the train a maximum deceleration of 2 m/s^2 . The top speed of the train is 75 m/s and each axle of the train carries a load of $15\,000 \text{ kg}$. (i) What is the retarding force on each axle as the train decelerates at its maximum rate and (ii) what is the initial power dissipated on each axle if the brakes are fully applied while the train is travelling at top speed?

(i) $F = ma$ (see page 34) Hence, $F = 15\,000 \times 2 = 30\,000 \text{ N}$

(ii) Power = work done per second (see page 56)

$$= 30\,000 \times 75 = 2.25 \times 10^6 \text{ J/s or } 2.25 \times 10^6 \text{ W}$$

MACHINES

7.1 INTRODUCTION

A machine is any device which enables mechanical work to be done more easily. Very often this results in a small applied force being amplified into a large one, for example when a screwdriver is used to 'lever' a lid off a tin. Sometimes a small movement is amplified into a large one, as in a bicycle when a small pedal movement results in a much larger movement of the bicycle along a road.

For any machine the principle of conservation of energy, stated in Chapter 6 Section 6.6, must apply. The source of energy for a moving bicycle is the chemical energy in the food eaten by the cyclist. Some of this is converted into work done against friction (due to moving parts and the air resistance) as the bicycle goes along.

7.2 AN EXPERIMENT WITH PULLEYS

The pulley system shown in Fig. 7.1 has a double pulley at the top and a single pulley at the bottom. The load is attached to the bottom single pulley. The effort required to raise the load takes the form of a weight hanger with slotted weights. The effort weight is gradually increased until, when given a small start, it will raise the load with constant velocity. The effort required to raise different loads in this way may be investigated. Table 7.1 shows a series of readings obtained in this way.

The third column in the table is the work done on the load, that is, the potential energy gained by the load when it is raised 1 m (load (N) $\times$ 1 (m)). The fourth column is the work done by the effort in raising the load. Notice that when the load comes up 1 m, each of three strings supporting the load shortens by 1 m and therefore the effort moves 3 m. Thus in this machine the effort moves three times as far as the load. The fourth column is therefore the effort multiplied by 3.

fig 7.1 a simple pulley system

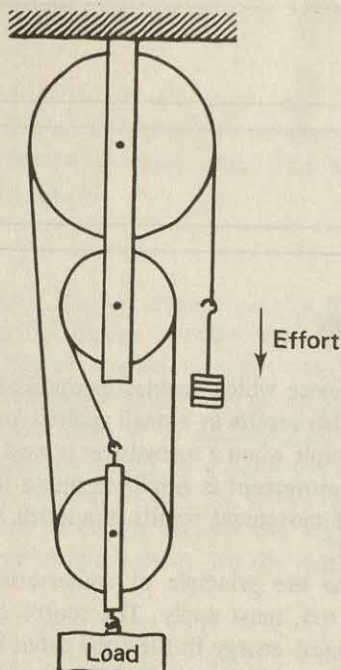


Table 7.1 results for the experiment illustrated in Fig. 7.1

Load (N)	Effort (N)	Work done on load in raising it 1 m (i.e. PE gained by load) (J)	Work done by effort (J) (= Effort (N) $\times$ 3 (m))	Efficiency = $\frac{\text{Work out}}{\text{Work in}}$
1.1	0.5	1.1	1.5	0.73
1.7	0.7	1.7	2.1	0.81
2.0	0.8	2.0	2.4	0.83
2.6	1.0	2.6	3.0	0.87
3.8	1.4	3.8	4.2	0.90
4.1	1.5	4.1	4.5	0.91
4.4	1.6	4.4	4.8	0.92
5.0	1.8	5.0	5.4	0.92

The efficiency of any machine is defined as follows:

$$\text{Efficiency} = \frac{\text{Work got out of machine}}{\text{Work put into machine}} \quad (7.1)$$

In this particular machine, the work got out of the machine is the potential energy gained by the load, and the work put into the machine is the work done by the effort. The results obtained for various loads are given in the fifth column of the table.

Notice that the efficiency increases with the load. In the absence of friction, and with a weightless bottom pulley the efficiency would be 100%. Because the effort always has to do work against friction and also work in lifting the bottom pulley block as well as the load, the efficiency of a pulley system is always less than 1.0, or less than 100%.

Of course the efficiency of any machine can never be more than 1.0, for we can never get more work out of a machine than what we put into it.

7.3 TWO DEFINITIONS FOR MACHINES

As stated above, for the pulley system illustrated in Fig. 7.1 the effort moves three times as far as the load moves. We say that the velocity ratio (VR) of the machine is 3 because the velocity ratio is defined by the equation

$$\text{Velocity ratio} = \frac{\text{Distance moved by effort}}{\text{Distance moved by load}} \quad (7.2)$$

Another quantity which is used in talking about a machine is the mechanical advantage (MA) defined by the equation

$$\text{Mechanical advantage} = \frac{\text{Load}}{\text{Effort}} \quad (7.3)$$

For the pulley system giving the results in Table 7.1, notice that when the load was reasonably large its value was nearly three times that of the effort. We say that the mechanical advantage was nearly 3. If the bottom pulley block were weightless, and if there were no friction in the system, the mechanical advantage would be exactly 3 and equal to the velocity ratio.

Although both the mechanical advantage and the efficiency of a machine depend on the friction in the machine, the velocity ratio depends only on the size and shape of the machine parts.

Two further points can be made about the use of pulleys in machines. Fig. 7.2 indicates that a pulley may be used to change the direction of a force needed to move a load. In Fig. 7.3, neglecting the weight of the pulley itself, the load is supported by two upward forces, one on each side of the

fig 7.2 *this pulley wheel changes the direction of the applied force*

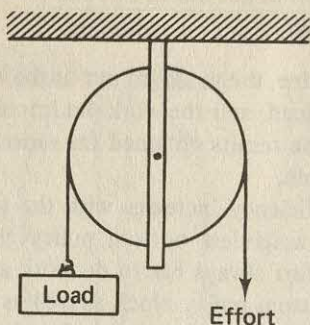
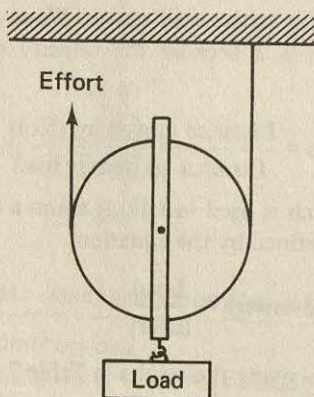


fig 7.3 *this pulley reduces the force needed to raise the load (with a weightless bottom block and no friction the effort would be half the load)*



pulley wheel, each equal to half the load. The force needed to raise the load is half the load (because there is a constant tension force throughout the string equal to the effort force). Both these features find application in machines using pulleys.

7.4 THE INCLINED PLANE

A very ancient device for raising loads (used by the Pharaohs) is the inclined plane, still to be seen on building sites, in the winding roads up mountain passes or the sloping planks sometimes used to get beer barrels

onto lorries. For the inclined plane

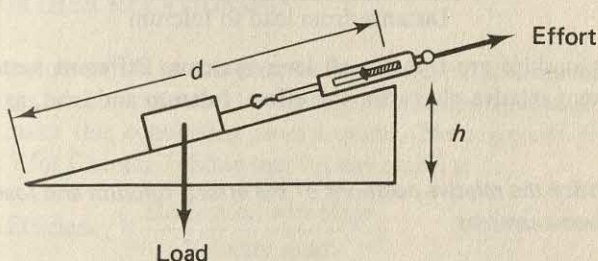
$$\text{Velocity ratio} = \frac{\text{Distance moved along the plane}}{\text{Height through which load is lifted}} = \frac{d}{h}$$

(see Fig. 7.4), and

$$\text{Mechanical advantage} = \frac{\text{Weight of object}}{\text{Force needed to move object up the plane}}$$

The mechanical advantage must be measured by determining the force necessary to move the object.

fig 7.4 *the inclined plane*

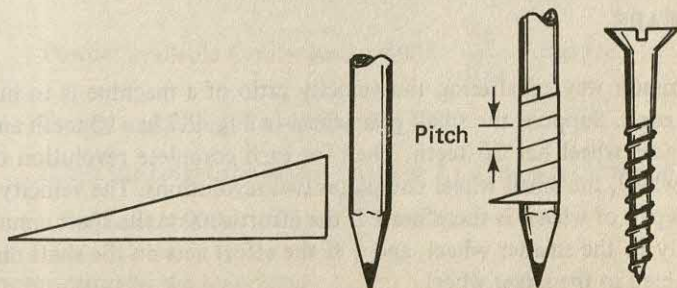


7.5 THE SCREW

A screw thread is a kind of inclined plane, spiralled round the axis of the screw (Fig. 7.5). As the screw is rotated once, it advances the distance between the threads, known as the pitch of the screw. So, for a screw

$$\text{Velocity ratio} = \frac{\text{Circumference of screw}}{\text{Pitch of screw}}$$

fig 7.5 *the screw is a kind of spiral inclined plane*



In general, machines with a large velocity ratio will tend to have a large mechanical advantage (see Section 7.8) and this applies to both the screw and the inclined plane.

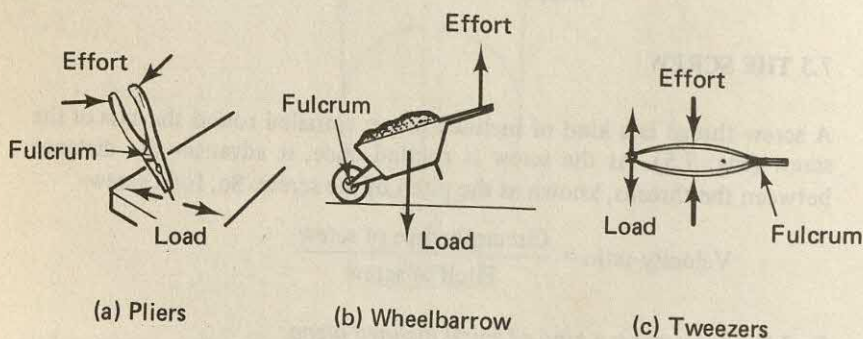
7.6 LEVERS

One kind of lever has already been mentioned at the beginning of Chapter 5. In the screwdriver example, the load is the force needed to move the tin lid and the effort is the force exerted on the screwdriver handle. The velocity ratio is given by

$$\begin{aligned}\text{Velocity ratio} &= \frac{\text{Distance moved by effort}}{\text{Distance moved by load}} \\ &= \frac{\text{Distance from effort to fulcrum}}{\text{Distance from load to fulcrum}}\end{aligned}$$

These relationships are true for all lever systems. Different systems can have different relative places for the effort, fulcrum and load, as Fig. 7.6 indicates.

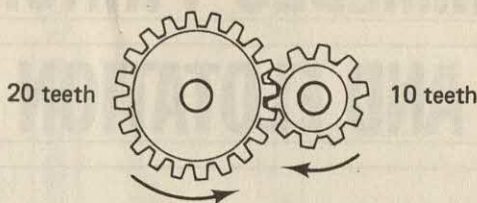
fig 7.6 notice the relative positions of the effort, fulcrum and load in the given examples



7.7 GEARS

A common way of altering the velocity ratio of a machine is to include some gears. Suppose the small gear wheel in Fig. 7.7 has 10 teeth and the large gear wheel has 20 teeth. Then for each complete revolution of the large wheel, the small wheel completes two revolutions. The velocity ratio of this pair of wheels is therefore 2 if the effort acts on the shaft connected directly to the smaller wheel, and $\frac{1}{2}$ if the effort acts on the shaft directly connected to the larger wheel.

fig 7.7 when the small wheel rotates twice, the large wheel completes just one revolution



In general, for gears

$$\text{Velocity ratio} = \frac{\text{Number of teeth on one gear wheel}}{\text{Number of teeth on other gear wheel}}$$

7.8 A FURTHER RELATIONSHIP

The discussion so far in this chapter has suggested that machines which have a large velocity ratio also have a large mechanical advantage. It is possible to make this connection more explicit. The argument indicated in question 7 for Chapter 7 shows that for any machine

$$\text{Efficiency} = \frac{\text{Mechanical advantage}}{\text{Velocity ratio}} \quad (7.4)$$

As the efficiency of any real machine must always be less than 1.0 (because some energy will always be lost due to friction), the mechanical advantage of any machine must always be less than its velocity ratio, a fact implied at the end of Section 7.3. For a machine which has a very high efficiency, the mechanical advantage will be nearly equal to the velocity ratio.

WORKED EXAMPLE

An electric motor draws energy from the mains at a rate of 600 W. If the motor is 50% efficient how far can the motor raise a load of 45 N in 3 s?

$$\text{Powder available to raise load} = 600 \times \frac{50}{100} = 300 \text{ J/s}$$

$$\text{Energy available in 3 s} = 300 \times 3 = 900 \text{ J}$$

$$\begin{aligned} \text{If } h \text{ is the height the load is raised in 3 s, then work done on load} \\ = 45 \times h = 900 \text{ J} \quad \therefore h = \frac{900}{45} = 20 \text{ m} \end{aligned}$$

The motor can raise the load 20 m

ARCHIMEDES' PRINCIPLE AND FLOTATION

8.1 INTRODUCTORY EXPERIMENT

Why is it that ships made of many tons of iron can float on water, while a piece of solid iron thrown into water will sink?

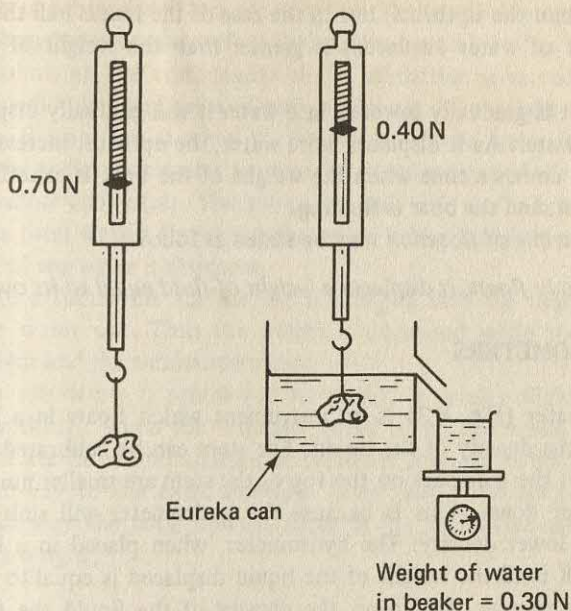
The following experiment will help you to answer this question. A stone is hung from a spring balance and its weight recorded. The stone, still attached to the spring balance, is then immersed in some liquid and the weight again recorded. When the stone is immersed in the liquid it appears to weigh less. The reason for this is that the liquid is pushing up on the stone. The upward force of a liquid on an immersed object is known as the *upthrust*. Table 8.1 shows some readings for this experiment when the stone is immersed first in water, then in paraffin and finally in salt solution. The apparent loss of weight (the upthrust) when the stone is immersed in water is $0.70 - 0.40 = 0.30$ N.

Table 8.1

	<i>Liquid</i>		
	<i>Water</i>	<i>Paraffin</i>	<i>Salt solution</i>
Weight of stone in air = 0.70 N			
Weight of stone in liquid (N)	0.40	0.46	0.34
Upthrust (N)	0.30	0.24	0.36
Weight of liquid displaced (N)	0.30	0.24	0.36

The experiment is conducted with a special kind of can called a eureka can. A eureka can has a spout on the side of it as shown in the diagram (Fig. 8.1). Such a can when full will always hold the same volume of liquid

fig 8.1 *the upthrust is equal to the weight of fluid which is displaced*



because the liquid will always settle at the same level. When the stone is immersed in the eureka can, the volume of liquid which comes out is exactly equal to the volume of the stone. Thus the liquid displaced by the stone comes out of the spout and may be collected in a beaker and weighed. The weight of displaced liquid is recorded in the last line of Table 8.1.

Study carefully the last two lines of the table. What conclusion can you come to? Archimedes' principle states that

When a body is completely or partially immersed in a fluid, the upthrust on it is equal to the weight of fluid it displaces

The reason for the different readings with different liquids is due to the different liquid densities (see question 3 for Chapter 8).

The word 'fluid' is used because Archimedes' principle applies to gases as well as liquids (see Section 8.5).

8.2 THE PRINCIPLE OF FLOTATION

Suppose that one day when you are swimming you dive under the water with a stone in one hand and a tennis ball in the other. If you release both

objects when you are under the water, what will happen? The stone will sink and the ball will rise. In the case of the stone, the weight of the stone is greater than the upthrust, but in the case of the tennis ball the upthrust (the weight of water displaced) is greater than the weight of the tennis ball.

If a boat is gradually lowered into water it will gradually displace more and more water. As it displaces more water, the upthrust increases. Eventually there comes a time when the weight of the boat is exactly equal to the upthrust, and the boat is floating.

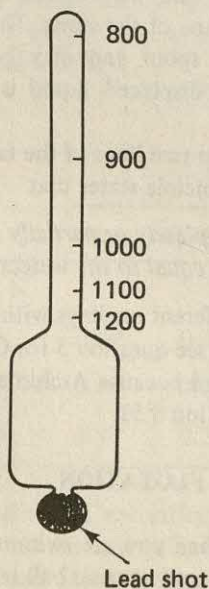
The *principle of flotation* may be stated as follows:

When a body floats, it displaces a weight of fluid equal to its own weight

8.3 HYDROMETERS

A hydrometer (Fig. 8.2) is an instrument which floats in a liquid and measures the density of the liquid. The stem can be calibrated in kg/m^3 . Notice that the markings on the top of the stem are smaller numbers than those lower down. This is because the hydrometer will sink farther in liquids of lower density. The hydrometer, when placed in a liquid, will always sink until the weight of the liquid displaced is equal to the weight of the hydrometer. The lower the density of the liquid the farther the hydrometer will have to sink until this occurs.

fig 8.2 a hydrometer



8.4 SUBMARINES

In the hulls of submarines there are very large tanks called buoyancy tanks. When a submarine is on the surface these tanks have a lot of air in them. In order to submerge, the commander of the submarine gives orders for the tanks to be flooded and water flows into them. The weight of the submarine on the surface is equal to the weight of sea water it displaces. Flooding the tanks with water increases the total weight of the submarine which therefore submerges. The submarine will sink to the bottom of the ocean if its total weight (including the water in the tanks) is greater than the weight of sea water it displaces.

To make a submarine surface, air is pumped into the tanks and this pushes the water out. Thus the weight is decreased while the upthrust stays constant and the submarine rises.

When a submarine is submerged its weight is usually slightly greater than the upthrust. When the required depth is reached small 'trimmers' fore and aft are tilted and additional forces act upwards on the submarine in a similar way to the extra sideways force caused by the rudder of a ship. In this way the total upward force in the submarine can be made to equal the weight.

8.5 THE BALLOON

The principle which explains how a submarine floats submerged in the sea also explains how a balloon floats in air.

If a balloon is filled with hydrogen, the upthrust on it (equal to the weight of air displaced) is greater than the weight of the balloon (fabric plus hydrogen). There is therefore a resultant upward force and the balloon rises.

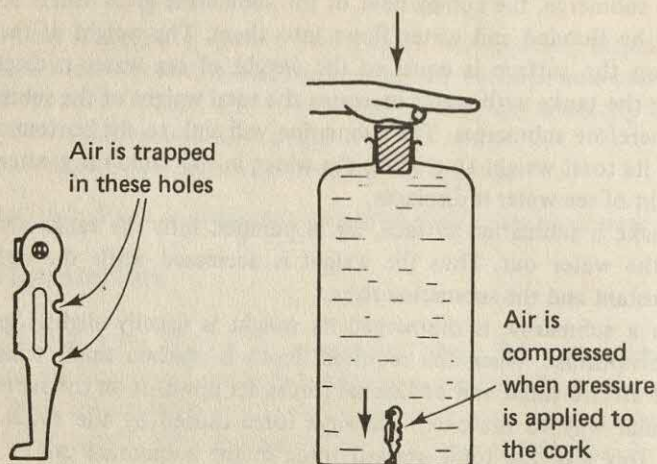
Gases other than hydrogen may be used. Hot air, for instance, has a density less than air at normal temperature. If a balloon is filled with hot air, the upthrust due to the displaced colder and denser air may be greater than the total weight of the balloon (fabric plus hot air). Under these circumstances the balloon will rise.

8.6 THE TOY DIVER

The toy diver shown in Fig. 8.3 sinks to the bottom of the bottle when pressure is applied to the cork and rises again when the pressure is released. The diver has a number of small holes in his body in which air is trapped. When the pressure is increased, water enters the holes. The weight of the diver is thus increased and he sinks. When the pressure on the cork is

released the water comes out of the holes, the weight of the diver is reduced and he rises.

fig 8.3 a toy diver



WORKED EXAMPLE

A research balloon has fabric of mass 100 g and a gas capacity of 2 m^3 . It is filled with hydrogen of density 0.085 kg/m^3 . What is the maximum weight of instruments that the balloon can lift? (Density of air = 1.22 kg/m^3).

$$\text{Mass of hydrogen in balloon} = 0.085 \times 2 = 0.17 \text{ kg}$$

$$\text{Mass of balloon plus fabric} = 0.1 + 0.17 = 0.27 \text{ kg}$$

$$\text{Total weight of balloon} = 0.27 \times 10 = 2.7 \text{ N}$$

$$\text{Mass of air displaced by balloon} = 1.22 \times 2 = 2.44 \text{ kg}$$

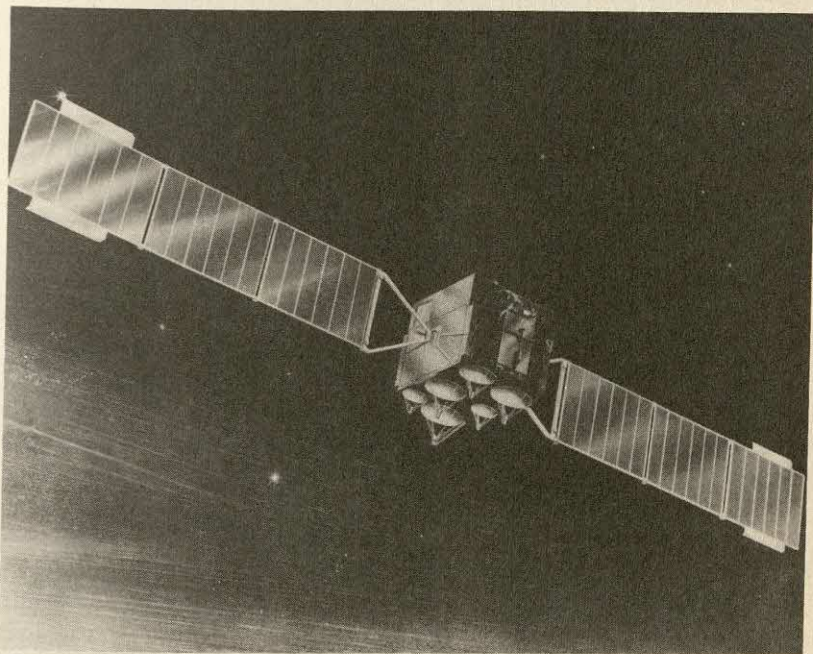
$$\text{Weight of air displaced by balloon} = \text{Upthrust} = 2.44 \times 10 = 24.4 \text{ N}$$

$$\begin{aligned} \text{Weight of instruments that can be lifted at ground level} \\ &= (\text{Upthrust}) - (\text{weight of balloon}) \\ &= 24.4 - 2.7 = 21.7 \text{ N} \end{aligned}$$

PART II

HEAT AND MOLECULAR STRUCTURE

European Space Agency's European communications satellite. Solar panels convert the Sun's rays into electrical energy (courtesy British Aerospace)



HEAT ENERGY

9.1 INTRODUCTION

If you laze on a beach in the sunshine you are aware of the heat energy from the Sun falling on your body as you get hotter and hotter! This experience is an example of a basic physical fact; heat energy supplied to a body causes its temperature to rise.

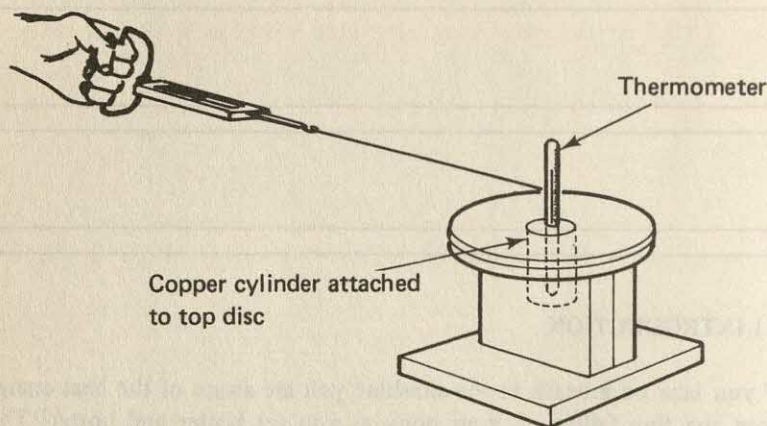
9.2 FRICTIONAL HEATING

There is another rather different way whereby a body may be heated. After a car has come down a steep mountain pass, the brakes are often quite hot to touch. We explain this by saying that the brakes have increased in temperature, and therefore gained in heat energy, because the car has lost potential energy (in coming down the mountain). The means by which the potential energy is converted into heat energy is the frictional force within the brakes. Whenever there are frictional forces between two surfaces as one surface moves across the other, work must be done against the frictional forces and heat energy will be produced as a result.

9.3 SPECIFIC HEAT CAPACITY

The ideas expressed in the previous section may be made more precise using the apparatus illustrated in Fig. 9.1. It consists of a copper cylinder clamped between two felt pads. A thermometer can be inserted into a hole in the middle of the cylinder. A wooden disc is fitted to the top of the cylinder and string may be wound round the circumference of this. The end of the string is attached to a spring balance so that when the spring balance is pulled the copper cylinder rotates. The friction between the copper cylinder and the felt pads produces heat energy and this causes a rise in temperature of the cylinder.

fig 9.1 apparatus with which the specific heat capacity of copper may be determined



Suppose 5 m of string are wound round the circumference of the disc and the spring balance is pulled so that it shows a steady force of 40 N. It is pulled with this steady force until it has moved 5 m. The temperature rise of the copper cylinder may be determined by reading the thermometer at the beginning and the end. A typical set of readings is shown below:

Mass of copper cylinder = 100 g = 0.1 kg

Rise in temperature = 5.1°C

Distance string was pulled = 5 m

Reading on spring balance = 40 N

The work done when 40 N moves through 5 m is $40 \times 5 \text{ J} = 200 \text{ J}$. Therefore

200 J cause 0.1 kg of copper to rise in temperature 5.1°C

$\frac{200}{0.1} \text{ J}$ cause 1.0 kg of copper to rise in temperature 5.1°C

$\frac{200}{0.1 \times 5.1} \text{ J}$ cause 1.0 kg of copper to rise in temperature 1.0°C

The quantity of heat energy required to raise the temperature of 1 kg of a substance through 1°C is called the specific heat capacity of the substance

So this experiment gives a value for the specific heat capacity of copper as

$$\frac{200}{0.1 \times 5.1} \text{ J/kg K}^* = 392 \text{ J/kg K}$$

The specific heat capacities of some substances are given in Table 9.1.

Table 9.1 *specific heat capacities (J/kg K)*

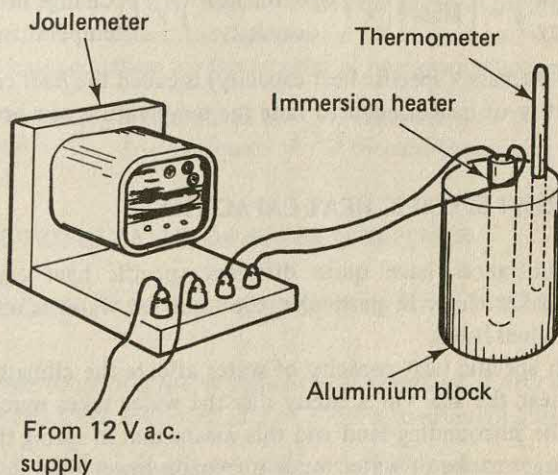
Aluminium	900	Ice	2100
Copper	380	Methylated	
Mercury	140	spirit	2500
Lead	140	Water	4200

9.4 SPECIFIC HEAT CAPACITY BY AN ELECTRICAL METHOD

A simpler, if less direct, method of determining specific heat capacity than the one just described, is one in which the energy is supplied electrically. An aluminium block has two holes bored in it, one in the centre in which a small immersion heater fits and another one in which a thermometer fits. The immersion heater is connected to a 12 volt a.c. supply via a joulemeter (Fig. 9.2).

The joulemeter is read and the initial temperature recorded. Then the electrical supply is switched on. When the temperature has risen about 5°C the supply is switched off and the joulemeter read again. A typical set of readings is shown overleaf:

fig 9.2 *apparatus for determining the specific heat capacity of aluminium electrically*



*An interval of one degree on the kelvin scale of temperature (K) is the same as an interval of one degree on the Celsius scale ($^{\circ}\text{C}$). Thus $392 \text{ J/Kg } ^{\circ}\text{C}$ and 392 J/kg K are identical (see also section 11.10 *et seq.*).

Mass of aluminium	= 1.0 kg
Initial temperature	= 20.0°C
Initial joulemeter reading	= 12 200 J
Final joulemeter reading	= 16 650 J
Final temperature	= 25.0°C

From these readings it follows that

1.0 kg of aluminium is heated through 5.0°C by
 $(16\,650 - 12\,200) \text{ J} = 4450 \text{ J}$

1.0 kg of aluminium is heated through 1.0°C by

$$\frac{4450}{5.0} \text{ J} = 890 \text{ J}$$

Hence the specific heat capacity of aluminium = 890 J/kg K.

9.5 A BASIC EQUATION

If a substance has a specific heat capacity of c J/kg, then

c J are required to raise 1 kg of the substance through 1 K

mc J are required to raise m kg of the substance through 1 K

mct J are required to raise m kg of the substance through t K

In words we say that

$$\left(\begin{array}{c} \text{Quantity of} \\ \text{heat energy} \end{array} \right) = \left(\begin{array}{c} \text{Mass} \end{array} \right) \times \left(\begin{array}{c} \text{Specific heat} \\ \text{capacity} \end{array} \right) \times \left(\begin{array}{c} \text{Change in} \\ \text{temperature} \end{array} \right)$$

The quantity (mass $\times$ specific heat capacity) is called the *heat capacity* and is the quantity of heat needed to raise the temperature of a body through 1°C.

9.6 DIFFERENT SPECIFIC HEAT CAPACITIES

Different substances have quite different specific heat capacities, as Table 9.1 makes clear. In particular, the value for water is much greater than other values listed.

The high specific heat capacity of water affects the climate near large lakes and near the sea. On a sunny day the water takes much longer to heat than the surrounding land and this means that in spring the temperature near large masses of water tends to remain lower than in other parts of the country. In autumn the temperature near large masses of water will be higher than that in other parts of the country.

9.7 THE METHOD OF MIXTURES

When a hot body and cold body come into contact they reach a common final temperature. If no heat energy is lost to the surroundings, the heat energy given out by the hotter body in cooling must equal the heat energy taken in by the cooler body in heating. This idea forms the basis of the method of mixtures, an example of which will now be given.

Suppose that 100 g of metal at 100°C is dropped into 50 g of liquid at 20°C and the final temperature of the mixture is 40°C . If the specific heat capacity of the liquid is 3000 J/kg K , what is the specific heat capacity of the metal?

Let the specific heat capacity of the metal be $c \text{ J/kg K}$. Using equation (9.1), 100 g (0.1 kg) of metal in cooling from 100°C to 40°C gives out

$$0.1 \times c \times (100 - 40) \text{ J}$$

50 g of liquid in heating up from 20°C to 40°C takes in

$$0.05 \times 3000 \times (40 - 20) \text{ J}$$

But the heat energy given out must equal the heat energy taken in, so

$$0.1 \times c \times (100 - 40) = 0.05 \times 3000 \times (40 - 20)$$

$$\therefore c = \frac{0.05 \times 3000 \times (40 - 20)}{0.1 \times (100 - 40)} = 500 \text{ J/kg K}$$

So the specific heat capacity of the metal is 500 J/kg K .

In this calculation we have neglected the heat taken in by the container of the liquid, often called the *calorimeter*. If that is to be taken into account, the basic equation for the transfer of heat energy becomes

$$\left(\begin{array}{c} \text{Heat given out} \\ \text{by metal} \end{array} \right) = \left(\begin{array}{c} \text{Heat taken in} \\ \text{by liquid} \end{array} \right) + \left(\begin{array}{c} \text{Heat taken in} \\ \text{by calorimeter} \end{array} \right)$$

9.8 THE CALORIFIC VALUE OF FUELS AND FOODS

When a fuel is burnt, its most important property is how much heat energy it gives out.

The calorific value of a fuel is defined as the amount of heat energy given out when unit mass (1 g or 1 kg) is completely burnt

A related idea is the calorific value of foods. The food we eat is oxidised in our bodies, the rate at which this is done (and the rate at which chemical energy is turned into heat energy) being just sufficient to keep our body

temperature constant when at rest. The energy being released by burning food just compensates for the energy lost from our skin to the surrounding cooler air. So the calorific value of a food, defined as the amount of heat energy released when unit mass of it is completely burnt in air, is an important quantity. You will, no doubt, have heard of Calories in connection with diet suggestions. The Calorie is a particular energy unit used by dieticians; 1 Calorie = 4200 J.

Some typical calorific values are given in Table 9.2.

Table 9.2 *approximate calorific values for some fuels and foods (J/g)*

<i>Fuels</i>		<i>Foods</i>	
Coal	28 000	Milk	3 000
Coal gas	28 000	Fats	40 000
Petrol	43 000	Proteins	17 000
Natural gas		Sugar	17 000
(North Sea gas)	56 000	Eggs	7 000
Wood	17 000	Cheese	20 000
		Potatoes	3 000
		Beef	15 000

WORKED EXAMPLE

A copper calorimeter of mass 100 g contained 150 g of oil at 20°C. A piece of copper of mass 40 g was quickly transferred from a hot oven to the calorimeter. The temperature of the oil rose by 10°C. What was the temperature of the oven? (The specific heat capacities of oil and copper are 1900 J/kg K and 380 J/kg K respectively.)

Let θ = temperature of oven. Then change in temperature of hot copper is $(\theta - 30)^\circ\text{C}$

$$\begin{aligned}
 \text{Heat given out by hot copper} &= \text{Heat taken up by oil} + \text{Heat taken up by calorimeter} \\
 \{0.04 \times 380 \times (\theta - 30)\} &= (0.15 \times 1900 \times 10) + (0.1 \times 380 \times 10) \\
 \therefore 15.2 \theta - 456 &= 2850 + 380
 \end{aligned}$$

$$\therefore \theta = \frac{3686}{15.2} = 242.5$$

EXPANSION

10.1 DEMONSTRATION OF EXPANSION

The experiments illustrated in Fig. 10.1 demonstrate the expansion of (a) a rod, (b) water and (c) air, when each is heated. When the rod is heated, the straw rotates. Now a small rotation of the needle results in a much larger movement of the end of the straw. If the length of the rod increases by an amount equal to the circumference of the needle, the straw will make one complete rotation. Solids expand very little on heating compared with liquids or gases and hence some arrangement like that shown in Fig. 10.1a is needed to magnify the movement.

When the flask in Fig. 10.1b is heated, the expansion of the water is seen as the water moves up the vertical tube. Gases expand very much more than liquids or solids when heated, and the warmth from the hand (Fig. 10.1c) is sufficient to cause the air to expand and bubble up through the water.

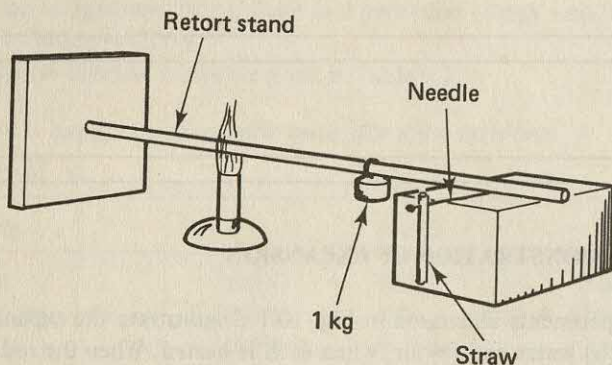
10.2 LARGE FORCES INVOLVED IN THE EXPANSION OF A SOLID

The very large forces involved in the expansion of a solid may be demonstrated using the apparatus shown in Fig. 10.2. The cast iron rod is inserted through the hole as shown and the knurled nut tightened so as to push the cast iron rod against the pillar at the opposite end of the apparatus. When the iron bar is heated with a bunsen burner, it expands and the tremendous force involved will snap the cast iron rod.

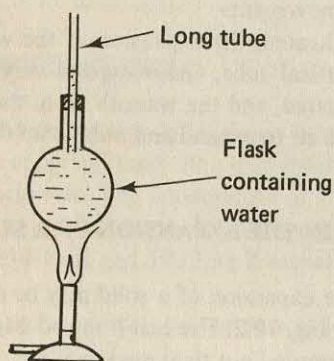
10.3 EXPANSION—FRIEND OR FOE?

The expansion of solids can be both an advantage and a disadvantage. The expansion is put to useful purpose when metal tyres are 'shrunk' on to wheels of railway locomotives. The tyres are fitted over the wheels when

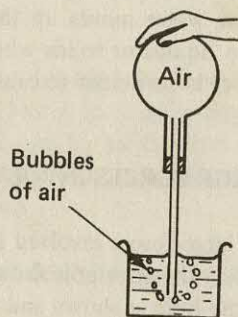
fig 10.1 (a) the expansion of the retort rod is small, but a small rotation of the needle produces a large movement of the straw end. (b) The liquid expands up the tube when heated. (c) Heat from the hand is sufficient to expand the air so that it bubbles out of the tube



(a)



(b)

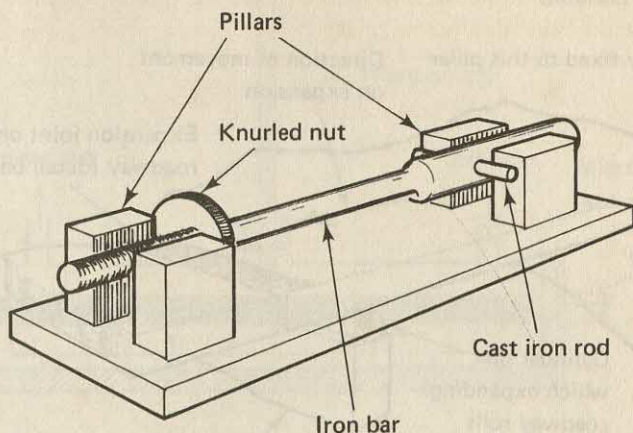


(c)

hot and when they cool they grip the wheel very tightly. The gear ring on the flywheel of a motor car engine is also fitted in this way.

Engineers must make allowance for expansion or else the large forces involved could cause their constructions to collapse. For example, in the Hammersmith flyover, London, the road surface is mounted on large rollers so that the whole of the surface of the flyover is able to move

fig 10.2 *breaking a cast-iron rod by the force resulting from the expansion of the iron bar*



when the temperature changes. The metal grid which runs across the fly-over near the midpoint allows the two parts of the flyover to expand towards each other without buckling (Fig. 10.3).

10.4 LINEAR EXPANSIVITY

The *linear expansivity* of a material is defined by the equation:

$$\text{Linear expansivity} = \frac{\text{Change in length}}{\text{Original length} \times \text{Temperature change}} \quad (10.1)$$

The unit of linear expansivity is just 'per degree' because the two lengths in the formula merely give a fraction without units. If the linear expansivity is known, then it is possible to calculate the amount of expansion of a given sample for a given rise in temperature, because all the quantities in equation (10.1) will be known except the 'change in length'. Table 10.1 gives some values of the linear expansivity for different materials.

To measure the expansivity for a sample in the form of a rod of material, all that is necessary is to measure the rod's original length at room temperature (this temperature must be measured), pass steam over it to heat it and measure the change in length as a result of the heating. A convenient apparatus is illustrated in Fig. 10.4. The initial length of the rod must be measured using a ruler. The micrometer screw gauge is used to measure the increase in length of the rod when heated. It is screwed up until it makes contact with the rod. The reading is recorded and the gauge unscrewed. The rod is heated and the screw gauge reading taken at regular intervals

fig 10.3 diagrams to indicate how expansion of a road bridge can be allowed for. (This method is used at the Hammersmith flyover, London)

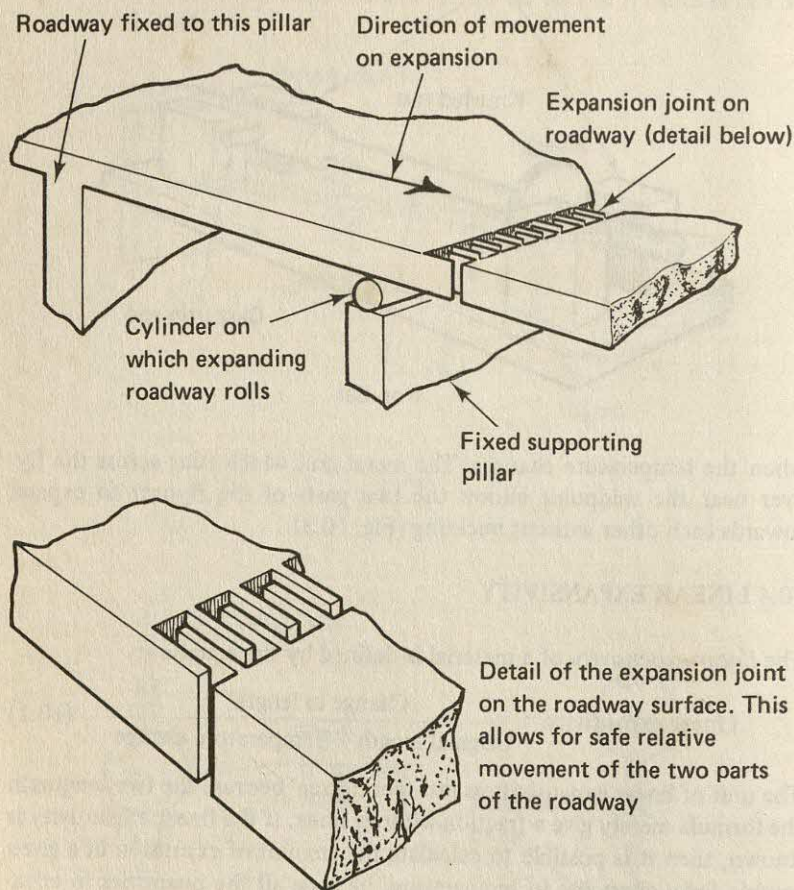
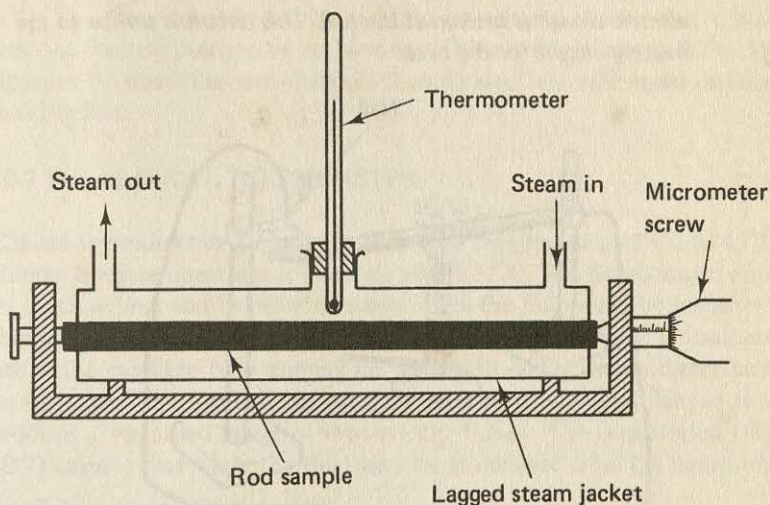


Table 10.1 *linear expansivities (per K)*

Aluminium	0.000 025	Platinum	0.000 009
Brass	0.000 019	Glass (ordinary)	0.000 009
Copper	0.000 017	Glass (Pyrex)	0.000 003
Steel	0.000 012	Invar	0.000 001

until it is constant. The difference between the initial and final reading on the screw gauge is the increase in the length of the rod. The linear expansivity is calculated using equation (10.1).

fig 10.4 apparatus for measuring the linear expansivity of a rod of material



10.5 BIMETALLIC STRIP

If two equal lengths of aluminium and iron are rivetted together to form a compound bar, when they are heated the bar will curve as illustrated in Fig. 10.5. The bar curves in the direction shown because aluminium expands more than iron when both are heated the same amount. Which way would the bar curve if the aluminium and iron were cooled below room temperature?

Fig. 10.6 illustrates the principle of an oven thermostat. When the bar is heated it curves downwards and no longer makes contact with the nail.

fig 10.5 a bimetallic strip (above) at a lower and (below) at a higher temperature

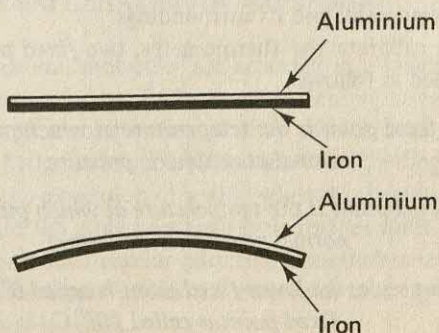
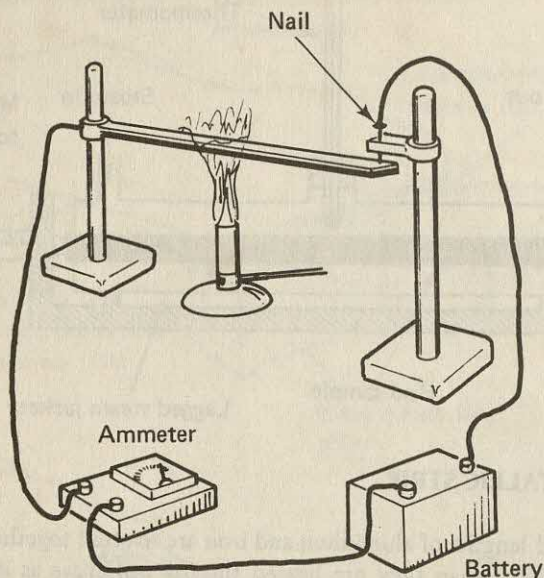


fig 10.6 *apparatus to illustrate the principle of an electric oven thermostat. When the bar temperature rises, the bar bends and the electric circuit is broken at the nail. This circuit is similar to the heating circuit for the oven*



The circuit is broken and the ammeter will no longer show a current. A similar device in an oven or an iron switches off the current when the element reaches a certain temperature.

10.6 CALIBRATION OF A MERCURY THERMOMETER

A mercury thermometer depends for its functioning on the fact that mercury expands more than glass, when heated. The level of mercury in the capillary tube above a bulb can therefore be used to indicate the temperature of the thermometer and its surroundings.

In order to calibrate the thermometer, two *fixed points* are required. These are defined as follows:

The lower fixed point is the temperature at which pure ice melts at normal atmospheric pressure

The upper fixed point is the temperature at which pure water boils at normal atmospheric pressure

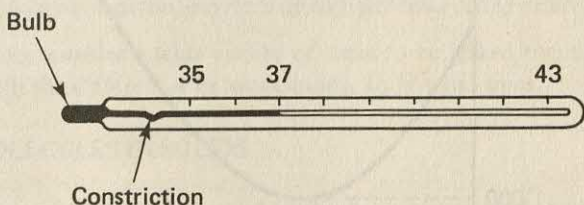
On the Celsius scale, the lower fixed point is called 0°C and the upper fixed point is called 100°C

To calibrate a thermometer it is first put into melting ice and a mark made on the glass to show the position of the mercury. This mark is labelled 0°C . The thermometer is then put into pure boiling water and a mark made at the new position of the mercury. This mark is labelled 100°C . The distance between the two marks is then divided into 100 equal divisions and labelled.

10.7 THE CLINICAL THERMOMETER

Clinical thermometers are usually calibrated over the range 35°C to 43°C . Human body temperature is normally about 37°C . The thermometer must be quick acting, and in order to achieve this the bulb must be made very thin and small. This means that the quantity of mercury used is small and hence the capillary tube running up the length of the thermometer must be very thin. The front of the capillary glass is shaped like a lens so as to produce a magnified image of the mercury thread. The constriction (Fig. 10.7) ensures that when the thermometer is removed from the mouth the

fig 10.7 a clinical thermometer



mercury in the capillary tube does not run back into the bulb. The mercury thread breaks at the constriction and the mercury in the capillary tube above the constriction remains in position so that the mouth temperature can still be read. A clinical thermometer is an example of a *maximum* thermometer.

10.8 MOLECULAR EXPLANATION OF EXPANSION

In a solid the atoms and molecules are arranged in a regular pattern and held together by strong forces. Each atom is vibrating to and fro about a fixed point and when a solid is heated they vibrate more violently and jostle each other a little further apart. This increased jostling and vibration helps us to explain the expansion of a solid when it is heated.

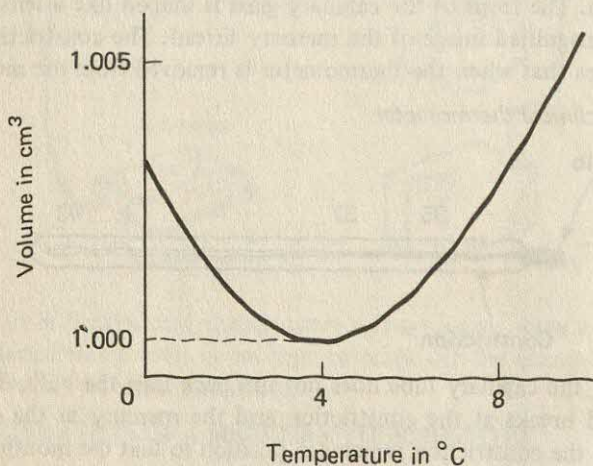
In the liquid state the molecules are a little further apart than in a solid and they do not form such regular patterns. Nonetheless increased movement at higher temperatures makes the molecules take up more room as the liquid temperature is raised.

In the gaseous state the molecules are much further apart and the attractive forces between them are now so small that the molecules move about freely. Gases exert a pressure on the walls of a container because the molecules collide with the walls. When the gas is heated the molecules move faster and hence the pressure increases. We discuss this further in Chapter 11, Sections 11.2 and 11.3.

10.9 THE ANOMALOUS EXPANSION OF WATER

Water is a most unusual liquid in that as its temperature *rises* from 0°C to 4°C its volume *contracts* (Fig. 10.8). Only above 4°C does the more normal expansion take place.

fig 10.8 a volume/temperature graph for water



This behaviour does not invalidate the statements in the previous section. The contraction of water as it is heated from 0°C to 4°C is caused by hydrogen bonds breaking down and causing a more compact structure. The change in volume due to this more compact structure more than compensates for the thermal expansion which simultaneously occurs.

PROPERTIES OF GASES AND MOLECULAR STRUCTURE

11.1 MOLECULAR THEORY OF MATTER

Many physical properties of matter can be discussed without regard to how matter is made up. But for a more complete discussion the nature of matter must be considered. The generally accepted theory is that:

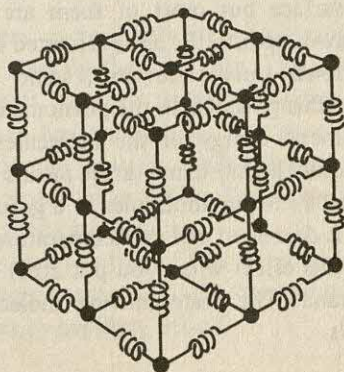
Matter consists of myriads of tiny particles called molecules

The theory enables a wide variety of facts to be linked together, and we begin with the differences between solids, liquids and gases.

11.2 MOLECULES IN SOLIDS

In solids, molecules are arranged in regular patterns and held together by strong attractive forces between them. Each molecule can vibrate about a fixed point. Fig. 11.1 shows a simple model of a solid. The molecules are

fig 11.1 *a simple model of a solid has balls to represent molecules joined by springs. The spring forces are equivalent to forces between the molecules.*



represented as balls joined together by springs, and if the model is shaken the balls vibrate backwards and forwards.

If the temperature of a solid is raised, the molecules are given more energy and (as this model suggests) they vibrate backwards and forwards with greater amplitude.

Notice that besides the strong attractive forces between the molecules there must also be repulsive forces which come into action if the molecules come very close together. It is these short-range repulsive forces which prevent a solid from being easily compressed.

In solids, the molecules are packed very tightly together, which is why the density of a solid is higher than that of the same substance as a liquid or gas.

11.3 MOLECULES IN LIQUIDS

When a solid is heated, the molecular vibrations become more violent. In the end the increased vibrations cause the regular pattern of molecules to be broken up. Each molecule becomes able to move away from its near neighbours. The material is melting. It is because of the relative freedom of movement of the molecules that a liquid can take the shape of its container.

The attractive forces between the molecules are not strong enough to give the liquid a definite shape, but nevertheless they are strong enough to hold the material together. The molecules are not free to wander off in any direction; they are held within the body of the liquid. At any given temperature the liquid therefore has a definite volume. This volume increases as the temperature increases, because at higher temperatures the molecules are moving about faster and do not fit into so small a volume as at lower temperatures.

At any particular temperature the molecules do not all move in the liquid at the same speed. The faster molecules are continually breaking through the liquid surface but most of them are attracted back again (rather as a stone thrown up into the air is attracted back to the surface by the Earth's gravitational force). However, some of the faster-moving molecules escape and this gives rise to the phenomenon known as *evaporation*. The average kinetic energy of the molecules is a measure of the temperature of the liquid; higher temperature means higher average kinetic energy. When the faster-moving molecules in a group escape, the average energy of those left is decreased and the temperature of the remainder has fallen. You can feel this effect when you put ether or some other volatile liquid onto your hand. The faster-moving molecules escape and the remaining liquid cools.

11.4 SURFACE TENSION

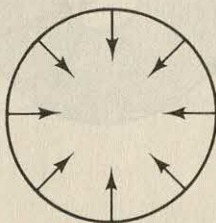
A needle which is placed on the surface of a piece of blotting paper floating on clean water will be found to float on the water when the blotting paper sinks. It is almost as if there were a skin on the water surface, on which the needle rests.

This property of a liquid results from the so-called cohesive forces between the liquid molecules. (Forces of attraction between like molecules are termed *cohesive*; forces of attraction between unlike molecules are termed *adhesive*.)

The cohesive forces between the molecules of water cause the water surface to behave somewhat as if it were a stretched membrane. The needle is prevented from piercing the surface of the water by the cohesive forces between the water molecules. It is rather as if the needle were resting on a stretched rubber membrane. For this reason the liquid property is termed *surface tension*.

Cohesive forces explain the spherical shape of liquid drops (Fig. 11.2).

fig 11.2 *cohesive forces on the surface molecules of a liquid drop pull them towards the body of the drop, until they are balanced by outward pressure forces. The result is a spherical drop*



They also explain the behaviour of a soap film in the demonstration illustrated in Fig. 11.3.

When a tube of very narrow bore (termed a capillary tube) is dipped into water, the water rises up the tube (Fig. 11.4). The reason for this is that the adhesive forces between water and glass are greater than the cohesive forces between water molecules. The glass above the water level attracts the water molecules upwards, while cohesive forces between water molecules draw more water up in the capillary tube behind the top molecules. The process is called capillary rise.

Blotting paper consists of a series of fine fibres with fine spaces between them. Any liquid making contact with the blotting paper will therefore be drawn into these spaces by capillary action.

fig 11.3 *when the soap film inside the central loose thread loop is pricked, the surface tension forces in the remaining areas pull the threads as shown*

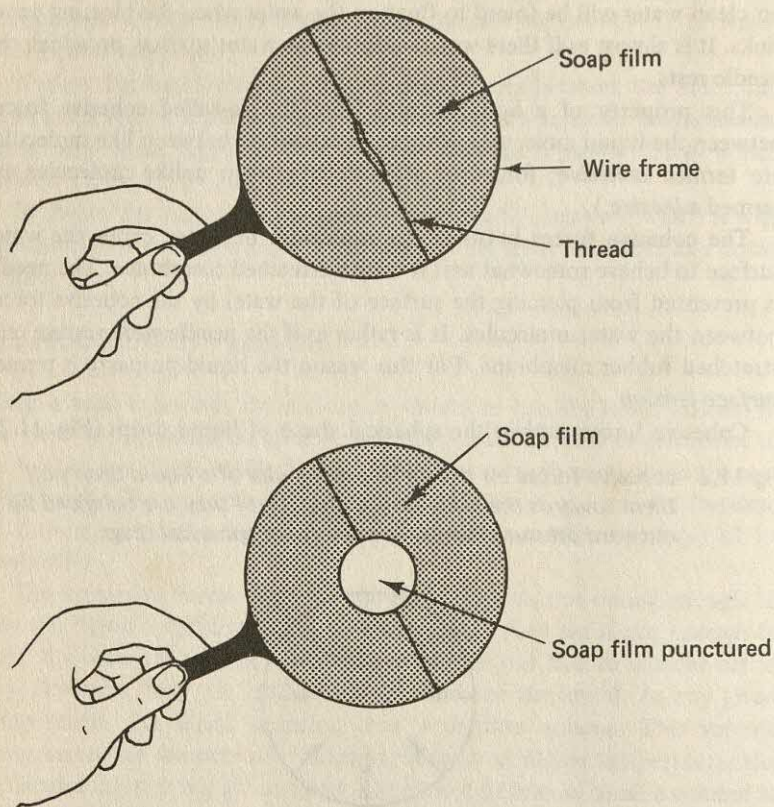
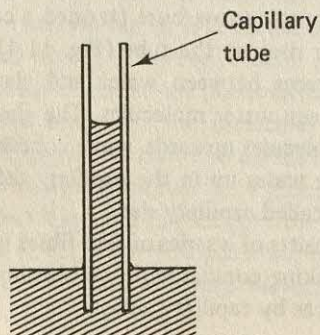


fig 11.4 *capillary rise up a narrow tube*



11.5 THE SIZE OF A MOLECULE

Some estimate can be made of the size of an oil molecule by performing the following experiment. Prepare a carefully cleaned water surface in, for example, a flat tray. Lightly dust the surface with lycopodium powder. Carefully let one drop of olive oil from a small dropper fall onto the centre of the water surface. It will rapidly spread out over the water surface, sweeping the powder in front of it.

Measure the diameter d of the oil drop before it leaves the dropper, by holding it close up against a millimetre scale. Measure also the diameter D of the patch of oil which it creates on the water surface.

The calculation of the thickness of the oil patch on the surface is as follows:

$$\text{Volume of oil drop} = \frac{4}{3}\pi r^3 = \frac{4}{3}\pi \left(\frac{d}{2}\right)^3 = \frac{\pi d^3}{6} \quad (11.1)$$

If we take the oil film thickness on the water to be h , then the volume of the film is

$$\pi R^2 h = \pi \left(\frac{D}{2}\right)^2 h = \frac{\pi D^2 h}{4} \quad (11.2)$$

But equations (11.1) and (11.2) are both expressions for the volume of the oil. Therefore

$$\frac{\pi d^3}{6} = \frac{\pi D^2 h}{4}$$

or

$$h = \frac{2d^3}{3D^2}$$

The value of h can therefore be calculated from the measurements. If we assume that the oil film spreads until the oil layer is just one molecule thick, the value of h gives an estimate of the size of one molecule. A typical value is about 10^{-9} m, a number almost inconceivably small. A thousand million such molecules placed side by side would only stretch for one metre.

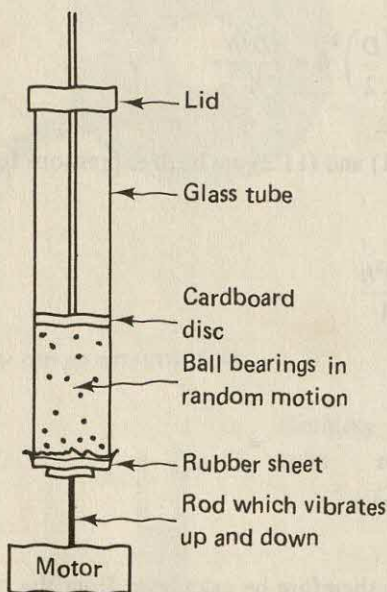
11.6 MOLECULES IN GASES

When sufficient heat energy has been added to a liquid, the molecules have enough speed to be able to break free from each other. The liquid has then become a gas. The molecules now move about freely throughout the complete space of their container, occasionally bumping into each other or the walls of the container. It is the collisions between the moving molecules and the walls of the container which cause the familiar gas pressure.

Fig. 11.5 illustrates an apparatus which demonstrates this model of gas behaviour. The rubber sheet at the bottom of the tube is agitated and moves up and down. The ball bearings resting on it are thrown about and hit the loose-fitting cardboard disc. The force of the balls hitting the disc causes it to rise until it reaches an equilibrium position.

This apparatus can demonstrate a further gas property (see Section 11.13).

fig 11.5 *an apparatus to demonstrate simple gas behaviour, in which ball bearings in random motion are the equivalent of gas molecules*



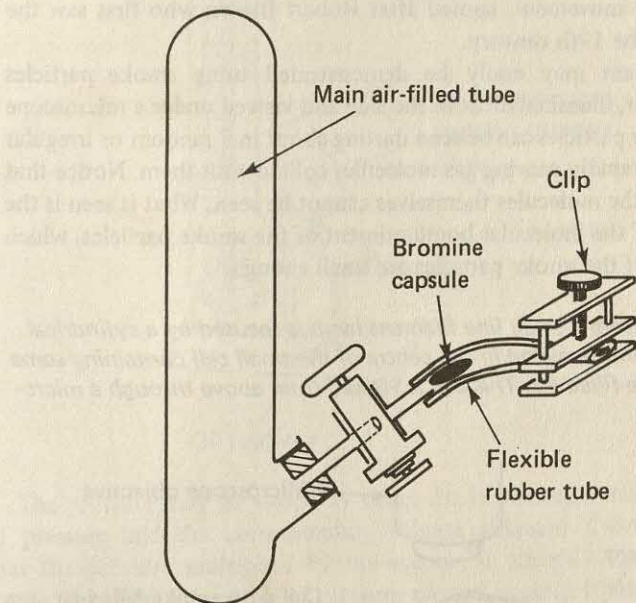
11.7 DIFFUSION

If some toast starts to burn in the kitchen, the smell quickly travels through the house. This indicates that the substance which smells moves from the

kitchen to other parts of the house. The movement of one kind of substance through a volume already occupied by something else is known as *diffusion*.

The most easily seen demonstration of gaseous diffusion involves the diffusion of bromine (a dense, brown-coloured gas) through air (much less dense). The apparatus, illustrated in Fig. 11.6, consists of a vertically mounted glass tube which near its base has a side tube containing a tap

fig 11.6 *an apparatus to demonstrate gaseous diffusion. Bromine is introduced from the side to the main tube and its colour gradually spreads up the main tube*



and ending in a closed rubber tube containing a sealed capsule of bromine.

The rubber tubing between the tap and the clip may be squeezed to break the capsule containing the bromine. When this happens, bromine is released into the space and as soon as the tap is opened it passes into the main tube. Its slow diffusion through the air already present in the tube may then be followed. If the tube is first evacuated, then when the capsule is broken the fast-moving bromine molecules immediately fill the tube.

Diffusion can also occur in liquids. If some blue copper sulphate solution is put into a tall jar and some distilled water added carefully on top, the blue colour will be subsequently found to spread gradually through the distilled water section.

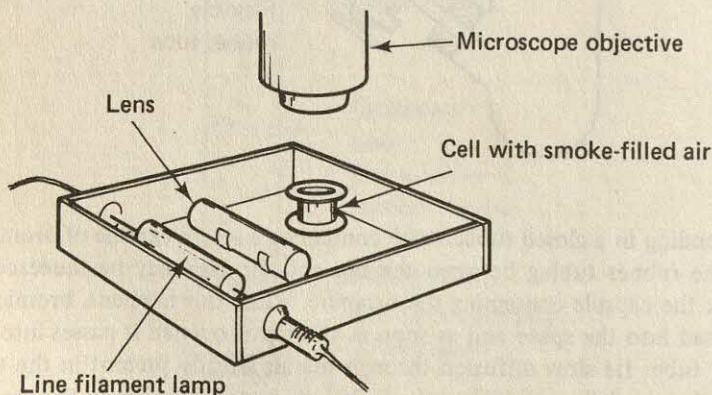
The process of diffusion in gases and liquids supports the molecular theory of matter. Molecules of one gas may pass easily between the molecules of any other gas, the rate of diffusion being limited by collisions between the molecules which are occurring all the time. There is less space between neighbouring molecules in a liquid, but the fact that molecules do not have fixed neighbours means that rearrangement continually takes place.

11.8 BROWNIAN MOVEMENT

More direct evidence for molecular movement in gases or liquids comes from Brownian movement, named after Robert Brown who first saw the effect early in the 19th century.

The movement may easily be demonstrated using smoke particles suspended in air, illuminated from the side and viewed under a microscope (Fig. 11.7). The particles can be seen darting about in a random or irregular manner as the rapidly moving gas molecules collide with them. Notice that the motion of the molecules themselves cannot be seen. What is seen is the consequence of the molecular bombardment of the smoke particles, which only shows up if the smoke particles are small enough.

fig 11.7 *light from a small line filament lamp is focused by a cylindrical lens onto a region in the centre of the small cell containing some smoke-filled air. This cell is viewed from above through a microscope*

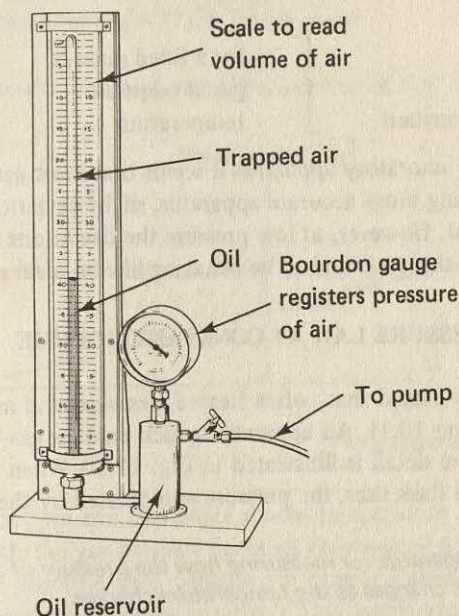


11.9 BOYLE'S LAW FOR A GAS AT CONSTANT TEMPERATURE

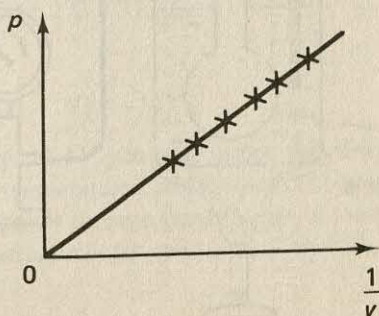
We now consider various properties of gases. We begin with the relationship between the pressure and volume of a fixed mass of air, using the apparatus

illustrated in Fig. 11.8. The volume of the trapped air is read on the graduated scale behind the air column; the pressure is read on the Bourdon gauge.

fig 11.8 *Boyle's law apparatus. (Redrawn from material kindly supplied by Philip Harris Ltd)*



The pressure may be varied by using the pump and a series of readings of pressure and the corresponding volume obtained. Calculation shows that the pressure multiplied by the volume is always constant. Alternatively, Fig. 11.9 is a graph of pressure plotted against $1/\text{volume}$, and this



gives a straight line graph through the origin. These results are consistent with Boyle's law which may be stated as follows:

For a fixed mass of gas at constant temperature, the pressure is inversely proportional to the volume

As an equation this result may be expressed as

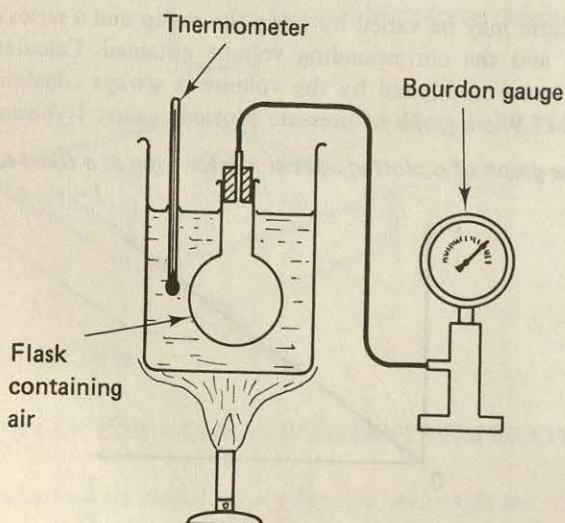
$$\left. \begin{array}{l} \text{or } p \propto \frac{1}{v} \\ pv = \text{constant} \end{array} \right\} \begin{array}{l} \text{for a fixed mass of} \\ \text{gas at constant} \\ \text{temperature} \end{array} \quad (11.3)$$

Using simple laboratory apparatus it seems that most gases obey Boyle's law exactly. Using more accurate apparatus, slight deviations from the law may be detected. However, at low pressure the deviations become less and in this situation the gas is said to be behaving like an ideal gas.

11.10 GAS PRESSURE LAW AT CONSTANT VOLUME

We have already stated that, when heated, gases expand more than liquids or solids (Section 10.1). An apparatus which enables gas behaviour to be explored in more detail is illustrated in Fig. 11.10. When the temperature of the air in the flask rises, the pressure as indicated by the Bourdon gauge

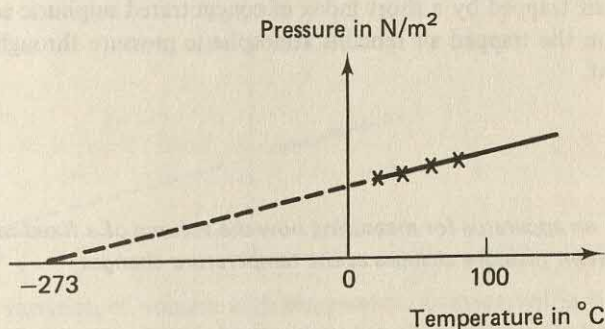
fig 11.10 *an apparatus for measuring how the pressure of a fixed volume of gas changes as the temperature changes*



goes up. Notice that the volume of the gas trapped in the flask and gauge stays constant.

The variation of pressure with temperature as measured by a mercury thermometer is illustrated by the graph in Fig. 11.11. Notice that the pressure falls as the temperature falls, and if the gas did not liquefy it is reasonable to suppose that the pressure would eventually become zero. This would happen at about -273°C for all gases.

fig 11.11 the results obtained with the apparatus of Fig. 11.10



Using this fact we can construct a new temperature scale. The temperature at which the gas pressure becomes zero we call the absolute zero. (It is called the absolute zero because we cannot imagine a situation in which the gas pressure becomes negative.) This temperature is called zero degrees Kelvin, abbreviated to 0 K. Temperature intervals on this new scale are chosen in such a way that a rise of 10°C is the same as a rise of 10 K. So the ice point, 0°C (being 273°C above the absolute zero), is 273 K.

In general, to convert a Celsius temperature into its Kelvin temperature add 273 to the Celsius temperature. To convert a Kelvin temperature to its Celsius temperature, subtract 273 from the Kelvin temperature. So

$$100^{\circ}\text{C} = 100 + 273 = 373 \text{ K}$$

and

$$100 \text{ K} = 100 - 273 = -173^{\circ}\text{C}$$

Looking again at Fig. 11.11, you may notice that if we alter the temperature scale to Kelvin temperatures, this is equivalent to moving the y-axis of the graph so that it passes through the intercept of the graph on the x-axis. Consequently the experimental results may now be expressed by the relationship

$$p \propto T \quad (11.4)$$

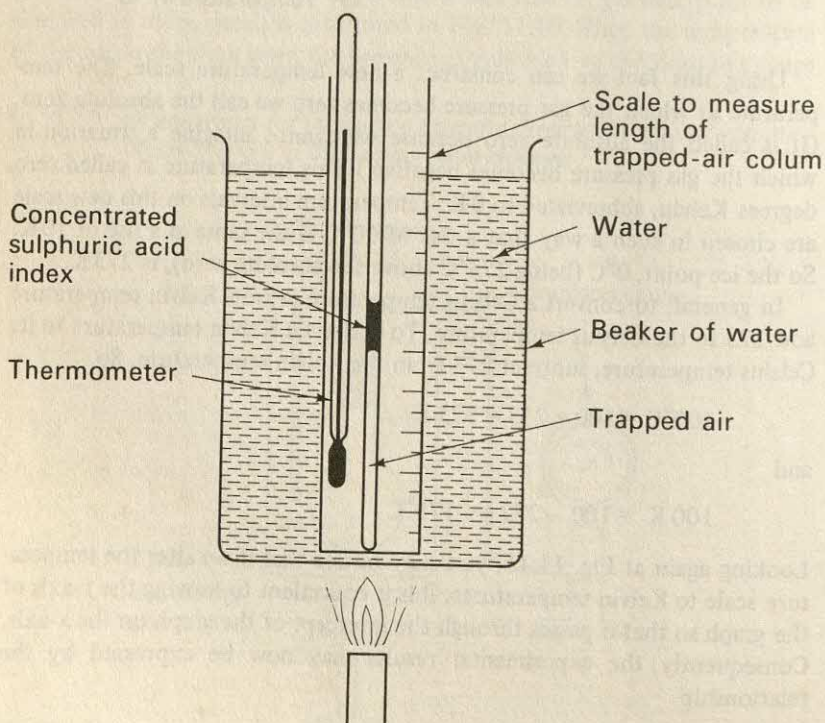
where p is the pressure, T is the absolute (Kelvin) temperature and v is kept constant, for a fixed mass of gas.

In other words, *for a fixed mass of gas at constant volume, the pressure is proportional to the absolute temperature.* This statement is sometimes known as the pressure law.

11.11 CHARLES' LAW FOR A GAS AT CONSTANT PRESSURE

For the next experiment the apparatus which is used is illustrated in Fig. 11.12. It consists of a wide-bore capillary tube, open at one end, containing some air trapped by a short index of concentrated sulphuric acid. The pressure on the trapped air remains atmospheric pressure throughout the experiment.

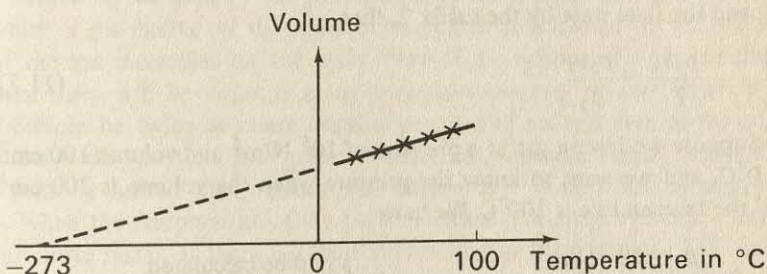
fig 11.12 an apparatus for measuring how the volume of a fixed mass of gas at constant pressure changes as the temperature changes



The sulphuric acid has two functions. It ensures that there is a fixed mass of air trapped in the tube; it also ensures that it is dry, free from water vapour which would modify its behaviour.

The capillary tube and a thermometer are immersed in water. The temperature of the water (and therefore of the air) varies and the resulting volume of the air measured by measuring the length of the capillary tube occupied by the air. The volume is proportional to the length of the tube occupied by the air.

fig 11.13 the results obtained with the apparatus of Fig. 11.12



The variation of volume with temperature as measured in this manner is illustrated in Fig. 11.13. Notice that the volume falls as the temperature falls and that, if the air did not first liquefy, its volume would become zero at about -273°C .

It is therefore convenient, as in the previous section, to express the results of this experiment using absolute (i.e. Kelvin) temperatures rather than Celsius temperatures. The results of the experiment indicate the following.

For a fixed mass of gas at constant pressure, the volume is proportional to the absolute temperature.

This statement is sometimes known as Charles' law.

In symbols the result may be expressed as

$$v \propto T \quad (11.5)$$

where v is the volume of the gas, T the absolute (Kelvin) temperature and the pressure, p , is kept constant.

11.12 THE IDEAL GAS EQUATION

Combining the results in equations (11.3), (11.4) and (11.5) we can write that the gas properties are in agreement with the equation

$$pv = kT \quad (11.6)$$

where k is a constant for a fixed mass of gas. Note that T must be the temperature expressed in Kelvin.

This equation leads to the concept of an ideal gas.

A gas which obeys the equation $p\nu = kT$ under all circumstances (i.e. at all pressures and temperatures) is called an ideal gas

In practice all real gases approximate at low pressures to the behaviour of an ideal gas.

It follows from the equation $p\nu = kT$ that if we represent the initial state of a gas by the suffixes 1, that is pressure p_1 , volume ν_1 and temperature T_1 , and the final state by the suffix 2, then

$$\frac{p_1\nu_1}{T_1} = \frac{p_2\nu_2}{T_2} \quad (11.7)$$

Suppose we have a gas at a pressure of 10^5 N/m^2 and volume 100 cm^3 at 0°C , and we want to know the pressure when the volume is 200 cm^3 and the temperature is 100°C . We have

$$p_1 = 10^5 \text{ N/m}^2$$

p_2 to be calculated

$$\nu_1 = 100 \text{ cm}^3 = 10^{-4} \text{ m}^3$$

$$\nu_2 = 200 \text{ cm}^3 = 2 \times 10^{-4} \text{ m}^3$$

$$T_1 = 273 \text{ K}$$

$$T_2 = 373 \text{ K}$$

Hence using

$$\frac{p_1\nu_1}{T_1} = \frac{p_2\nu_2}{T_2}$$

we have

$$\frac{10^5 \times 10^{-4}}{273} = \frac{p_2 \times 2.0 \times 10^{-4}}{373}$$

Therefore

$$p_2 = 6.83 \times 10^4 \text{ N/m}^2$$

11.13 THE KINETIC THEORY FOR GASES

In Section 11.6 we discussed the picture of a gas consisting of an assembly of molecules in rapid and random motion. The picture is sometimes referred to as the *kinetic theory of gases*.

Is the theory consistent with Boyle's law? An approximate answer to this comes from a more detailed use of the apparatus illustrated in Fig. 11.5. With the apparatus switched on, the disc settles out at an equilibrium position. The 'gas' volume may be measured in arbitrary units on a scale

alongside the tube. The gas 'pressure' may be altered by loading the disc with extra mass. When that happens the volume of the 'gas' of ball bearings also changes.

It is found that as the mass on the disc is changed, so that the effective pressure changes, the volume occupied by the ball bearings in equilibrium also changes in such a way that the product pV remains approximately constant. This model therefore shows that the moving ball bearings approximately obey the Boyle's law equation. Boyle's law is, in fact, in accord with the kinetic theory.

This may be seen by the following alternative argument. The pressure which a gas exerts on the walls of its container is caused by the impact of the gas molecules on the walls. Now if the volume of a gas is halved then there will be twice as many molecules per unit volume. There will therefore be twice as many impacts per second on unit area of the container walls, so that the pressure will have doubled. Hence, when the volume is halved the pressure will double and Boyle's law is obeyed.

When the temperature of the gas is increased, the molecules gain energy and move faster. If there is a fixed volume this increase in speed will cause an increase in the number of impacts per second and also a greater force on the wall for each impact. Consequently, the gas pressure will go up. This is in agreement with the experimental results obtained with the apparatus of Fig. 11.10.

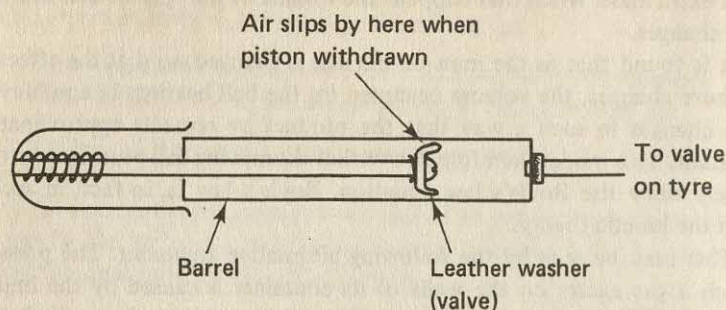
The kinetic theory of gases is, in fact, consistent with the gas properties described in this chapter. At a more advanced level, the ideal gas equation can be derived by applying the laws of motion to simple model molecules, and this affords a very important reason for belief in the theory.

Just as an increase in temperature means an increase in the average molecular speed, so a decrease in temperature means a decrease in the average speed. The absolute zero therefore represents the temperature at which the molecules have their lowest possible kinetic energy. On the other hand there seems to be no upper limit of temperature. The more kinetic energy the molecules have the higher is the temperature necessary to describe their condition.

11.14 THE BICYCLE PUMP

When the piston (Fig. 11.14) is pushed in, the air between the leather washer and the tyre valve is compressed. The greater pressure resulting from the compression of the air forces the edge of the leather washer against the side of the barrel forming an airtight seal. When the pressure of the compressed air exceeds the pressure of the air in the tyre, air is forced past the tyre valve into the tyre. When the piston is withdrawn the valve in the tyre prevents any air escaping from the tyre and the pressure

fig 11.14 a bicycle pump



in the space between the tyre valve and the leather washer falls. The edge of the leather washer is no longer forced against the side of the barrel and the atmospheric pressure forces air past the leather washer into the bottom part of the barrel.

WORKED EXAMPLE

A bubble of air doubles its volume in rising from the bottom to the surface of a lake. If the atmospheric pressure is 76 cm Hg and the density of mercury is $13\,600\text{ kg/m}^3$ what is the depth of the lake? Assume the temperature at the bottom and the surface of the lake are the same.

As the bubble rises its volume doubles, therefore the pressure of air inside it has halved (Boyle's law). So the pressure at the bottom of the lake is twice its value at the surface. Therefore the added pressure caused by the water is equal to the pressure of 76 cm Hg. Using equation 2.2 (page 17)

$$(10 \times h \times \rho)_{\text{water}} = (10 \times h \times \rho)_{\text{mercury}}$$

$$\therefore (h \times \rho)_{\text{water}} = (h \times \rho)_{\text{mercury}}$$

$$\therefore 1000 h_{\text{water}} = 13\,600 \times 0.76$$

$$\therefore h_{\text{water}} = \frac{13\,600 \times 0.76}{1000} = 10.3\text{ m}$$

$$\therefore \text{Depth of lake} = 10.3\text{ m}$$

THE TRANSFER OF HEAT ENERGY

12.1 INTRODUCTION

We are all familiar with the fact that heat energy can travel from one place to another. We can, for example, on sunny days feel the warmth from the Sun's rays which have travelled millions of miles across space. In the winter, houses are kept warm by heat energy which travels from fires and radiators to other parts of the buildings.

This chapter deals with the three different methods by which heat energy can travel from place to place, namely by *conduction*, *convection* or *radiation*.

12.2 CONDUCTION

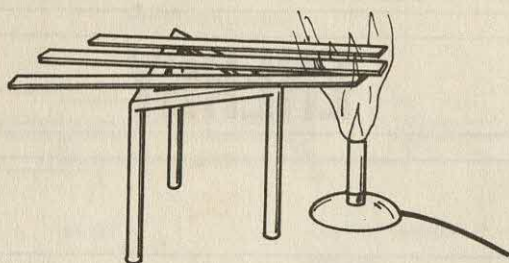
Conduction is the flow of heat energy through a body which is not at a uniform temperature, the energy flowing from places of higher temperature to places of lower temperature without the body as a whole moving

The flow of heat energy through the bottom of a saucepan (from the outside heat source to the inside contents) and the flow of heat energy along a teaspoon, one end of which is in a hot liquid, are examples of conduction.

Some substances conduct heat energy better than others. Fig. 12.1 shows an experiment to demonstrate this. The rods are made of different materials and one end of each rod is heated in a bunsen flame. After a little while, if you touch the cold end of each rod and slide your finger carefully towards the hot end, you will discover that some of the rods are conducting heat energy to their cold ends much better than are others.

Saucepans are made of metal because metals are good conductors of heat energy. Handles of cooking utensils, on the other hand, are made

fig 12.1 *an experiment to compare the conductivities of different materials*



from bad conductors so that they are not too hot to hold when the utensil is being used.

The experiment in Fig. 12.2 demonstrates that water is a bad conductor of heat energy. The ice is held in position at the bottom of the tube by a piece of wire gauze. The water at the top can be made to boil while ice remains at the bottom of the tube. The good conductivity of copper may be demonstrated by using a copper gauze as illustrated in Fig. 12.3. The gauze is held above the bunsen burner. If the gas is ignited above the gauze, the gas below the gauze does not ignite. This is because the gauze is a good conductor of heat energy and quickly conducts energy away

fig 12.2 *ice remains at the bottom of the test tube while water boils at the top, showing that water is a bad conductor of heat*

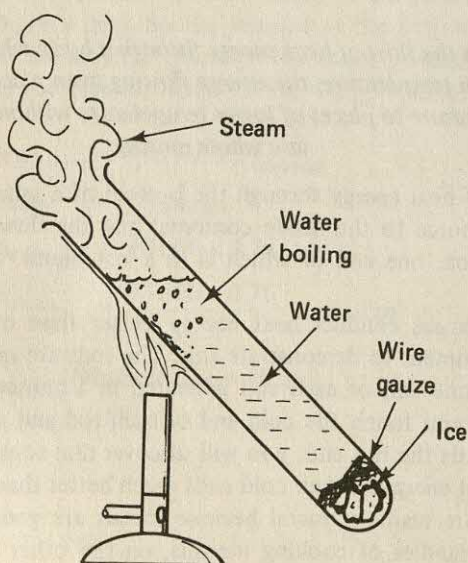
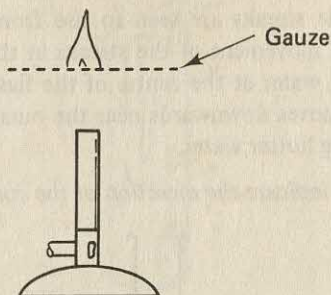


fig 12.3 *if the gas is lit above the gauze, the gas below the gauze does not ignite*



from the flame. The gas below the gauze therefore does not get hot enough to ignite.

Air is a bad conductor of heat energy and it is because of this property that cellular blankets, string vests, and fur coats are effective in keeping people warm. They all contain trapped air and are therefore bad conductors of heat energy.

12.3 AN EXPLANATION OF CONDUCTION

The conduction of heat energy along a rod from the hotter to the colder end can be explained using the molecular picture of matter discussed in the previous chapter. The molecules at the hot end are vibrating most vigorously and jostling the other molecules near to them. The heat energy is thus passed along the rod by interaction between the more vigorous molecules and the less vigorous ones. Metals are good conductors of heat energy (as well as being good conductors of electricity) for another quite different reason. Any metal contains large numbers of 'free' electrons which wander about within the metal rather like gas molecules inside a gas container. At the hot end of a rod of metal, the electrons have a higher kinetic energy than at the cold end. This kinetic energy is directly transferred to the colder parts of the metal rod. Because of the very large number of free electrons in a metal, and because of the high speed with which they move, heat energy is conducted very rapidly within a metal.

12.4 CONVECTION

Convection is the transfer of heat energy by the circulation of a fluid due to temperature differences within it

Figs 12.4, 12.5 and 12.6 illustrate experiments which demonstrate convection.

In Fig. 12.4 a tiny crystal of potassium permanganate is dropped to the bottom of a flask filled with water. When the bottom of the flask is gently heated, purple streaks are seen to rise from the slowly dissolving crystal. The upward movement of the streaks in the liquid indicates that the warm, less dense, water at the centre of the flask is rising. The colder, more dense, water moves downwards near the outside of the flask to take the place of the rising hotter water.

fig 12.4 *the arrows indicate the direction of the convection currents*

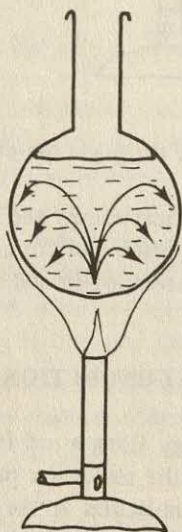
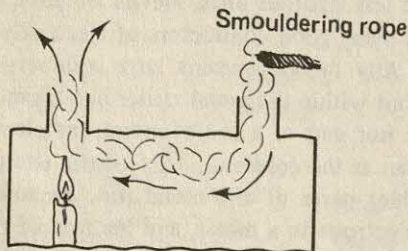
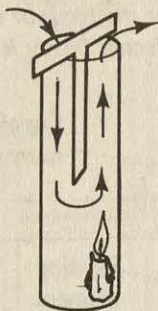


fig 12.5 *the arrows indicate the direction of the convection currents*



The apparatus shown in Fig. 12.5 consists of a closed box with two chimneys. A burning candle is placed under one chimney and a piece of smouldering rope held over the other. The movement of the smoke shows that the hot air is rising above the candle and cold air is coming down from the other chimney to take its place.

fig 12.6 *with the T-piece present, convection currents, indicated by the arrows, keep the candle alight. With the T-piece removed there is no circulation of fresh air past the flame and the candle goes out*



A very striking demonstration of convection currents in air may be given by using the apparatus shown in Fig. 12.6. In the absence of the T-piece, the burning candle goes out because there is not enough oxygen supply to keep it burning. When the T-piece is in place, the candle continues to burn because of the convection currents set up. Why do you think it works better if the T-piece is positioned so that it is not central but just to one side of the flame?

12.5 RADIATION

The heat energy which reaches us from the Sun cannot have reached us by either convection or conduction. Neither convection nor conduction can take place in a vacuum, and yet the heat from the Sun does reach the Earth. This third method by which heat energy can be transferred from one place to another is called radiation. If heat energy is manifested as molecular motion, it is clear that what travels through the space between the Sun and the Earth is not heat. It must be some form of energy since it becomes heat energy on reaching the Earth. The fact that, during an eclipse of the Sun, light and heat reaching the Earth are both cut off at the same time suggests that the heat energy which comes to us from the Sun is an electromagnetic radiation similar to light. This is in fact the case, so that

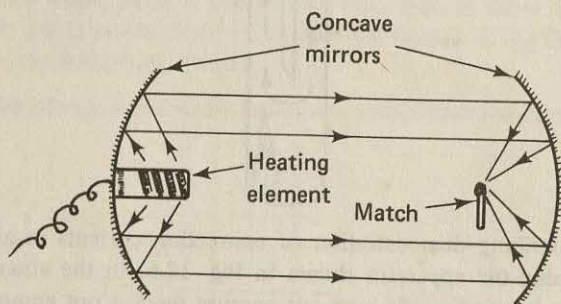
Radiation is the transfer of heat energy from one place to another by means of electromagnetic waves

12.6 REFLECTION AND REFRACTION OF RADIATION

Most people will at some stage have set fire to a piece of paper by using a magnifying glass to focus the Sun's rays on the paper. The radiation and the light are both refracted by the lens.

Fig. 12.7 shows how the heat from a heating element may be focused by concave reflectors onto the head of a match. The heat energy is concentrated at the match head sufficiently for the match to light.

fig 12.7 *the match is placed at the focus of the right-hand mirror and the radiation from the electric heater is concentrated enough to light the match*



12.7 EMITTERS AND ABSORBERS OF RADIATION

The emission of radiation may be investigated using the apparatus shown in Fig. 12.8. A copper sheet which is shiny on one side and blackened on the other is heated with bunsen burners. The sheet is then turned into a vertical plane and by placing the face first on one side of the sheet and then on the other the heat emitted by the blackened and shiny surfaces can be compared. The blackened surface is a much better emitter of radiation than the shiny surface.

The absorption of radiation may be investigated by using a heating element and a screen with a hole at its centre, as shown in Fig. 12.9. The absorption properties of glass may be investigated by putting your cheek close to the hole on the opposite side to the heating element and interposing a sheet of glass between the hole and your cheek.

Next hold the back of your hand near to the hole and repeat this with a very thin sheet of aluminium foil on the back of your hand. The aluminium foil may be fixed to the back of the hand by licking the foil before placing it on the hand. Finally coat the foil with a mixture of lampblack and methylated spirit and again hold your hand near to the hole. By these and similar experiments the following three conclusions may be drawn:

Shiny or white surfaces are not as good absorbers of radiation as black or dark surfaces

Polished or smooth surfaces are not as good absorbers of radiation as rough surfaces

fig 12.8 *the copper sheet is polished on one side and blackened on the other. The face can sense which side radiates more energy*

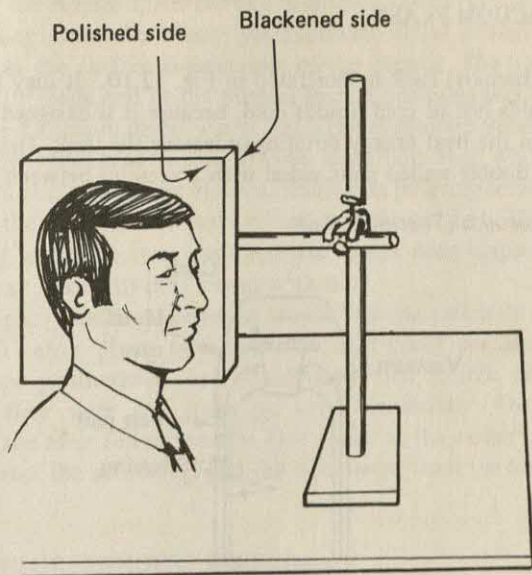
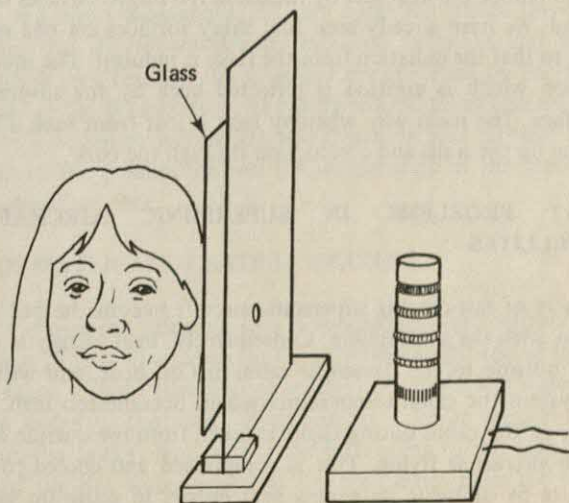


fig 12.9 *when the glass sheet is interposed between the heater and the cheek, the face temperature immediately goes down*

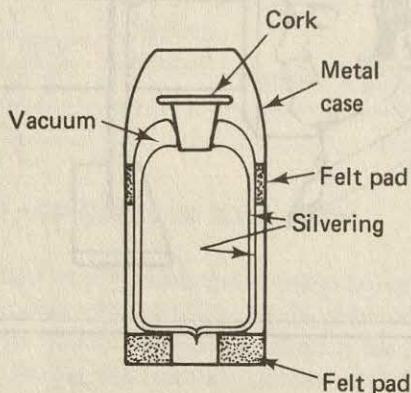


Black, rough surfaces are good emitters of radiation but light, shiny surfaces are poor emitters of radiation

12.8 THE VACUUM FLASK

A vacuum (Thermos) flask is illustrated in Fig. 12.10. It may be used to keep hot liquids hot or cold liquids cold, because it is designed to reduce to a minimum the heat energy entering or leaving the flask. The container consists of a double-walled glass vessel with a vacuum between the walls.

fig 12.10 a vacuum (Thermos) flask



Since conduction and convection cannot take place in a vacuum, heat energy cannot be lost from the container walls by either of these means. In order to reduce the heat lost by radiation the inside surfaces of the glass are silvered. We have already seen that shiny surfaces are bad emitters of radiation, so that the radiation from the flask is reduced. The small amount of radiation which is emitted is reflected back by the silvering on the other surface. The main way whereby heat is lost from such a flask is by conduction up the walls and conduction through the cork.

12.9 HEAT PROBLEMS IN SUPERSONIC AIRCRAFT AND SATELLITES

The surfaces of fast-moving supersonic aircraft become heated as a result of friction with the atmosphere. Consequently, heat energy is conducted from the outside to the passenger cabin in Concorde, and without some cooling system the cabin temperature would become too high. The fresh air for use in the cabin during flight is taken from the outside air through which the aircraft is flying. This is compressed and cooled to a suitable temperature by allowing its excess heat energy to warm up some of the

fuel in the fuel tanks. (The fuel temperature therefore gradually rises, from about 15°C to about 75°C during supersonic cruise.)

When the Apollo space capsule re-entered the Earth's atmosphere after the journey from the Moon, the frictional forces caused a tremendous increase in the surface temperature of the capsule. The blunt end of the capsule (the back end in the original launch position) was covered with a badly conducting material which was a honeycombed glass resin construction, called the heat shield. It was important that when the capsule re-entered the atmosphere the heat shield was pointing forwards, so that it and not the unprotected parts of the capsule got hot. The intense heat generated made the resin melt and the liquid resin dripped off and was swept away, taking its heat energy with it.

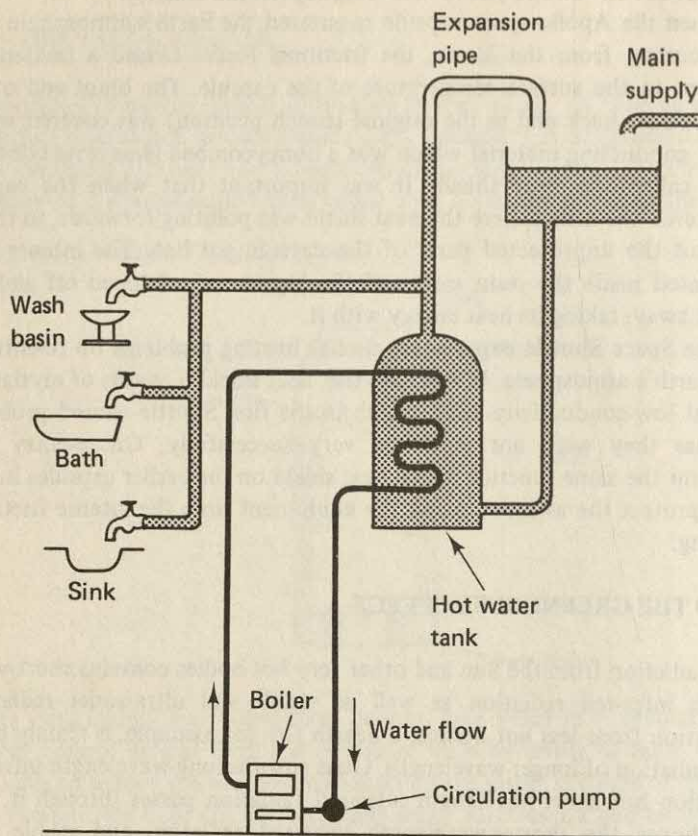
The Space Shuttle experiences similar heating problems on re-entry to the Earth's atmosphere. In this case the 'heat shield' consists of myriads of special low-conductivity tiles, which in the first Shuttle caused problems because they were not stuck on very successfully. On re-entry they perform the same function as the heat shield on the earlier capsules in that they protect the astronauts and the equipment from the intense frictional heating.

12.10 THE GREENHOUSE EFFECT

The radiation from the Sun and other very hot bodies contains short-wavelength infra-red radiation as well as visible and ultra-violet radiation. Radiation from less hot bodies, a hearth fire for example, is mainly infra-red radiation of longer wavelength. Glass absorbs long-wavelength infra-red radiation but short-wavelength infra-red radiation passes through it. In a greenhouse the shorter-wavelength infra-red radiation and visible light from the Sun pass through the glass and are absorbed by the soil and the plants, raising the temperature. The infra-red radiation emitted by the soil and plants is of longer wavelength and does not pass through the glass. It is trapped in the greenhouse and the temperature of the greenhouse and contents therefore rises.

12.11 HOT WATER AND CENTRAL HEATING

Fig. 12.11 shows the principle of a domestic hot water system. A convection current of hot water rises up the pipe leaving the top of the boiler. It passes through a spiral (the heat exchange coil) in the hot water tank, and heats the water in the tank. It returns at a lower temperature to the bottom of the boiler. Notice that the pipe going to the taps leaves from the top of the hot water tank and the cold water from the supply tank enters at the bottom of the hot water tank.

fig 12.11 *a simple domestic hot water supply system*

If a central heating system is included, a separate pipe leaves the top of the boiler taking water to the radiators and returns to the bottom of the boiler.

There are two main ways in which heat loss to the atmosphere from houses can be reduced. These are by filling cavity walls with a foam and putting an insulating material like fibreglass in the roof. Foam and fibreglass are bad conductors of heat and their presence minimises heat loss by convection currents in the air. Double glazing is also sometimes used. A layer of air is trapped between the two layers of glass and the poor conductivity of air minimises the heat loss.

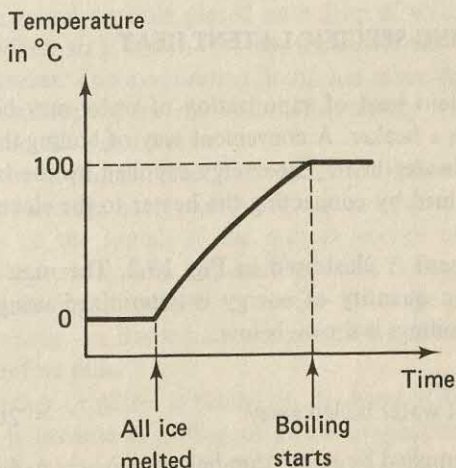
CHANGE OF STATE

13.1 LATENT HEAT

If some ice cubes from a refrigerator are put in a beaker of water and the mixture stirred, the temperature will be 0°C . If the beaker is now put on a tripod and heated very gently with a bunsen burner it will be found that the temperature remains at 0°C until all the ice has melted. The temperature of the water then begins to rise and goes on rising until it reaches 100°C . Heat energy is still being supplied to the beaker from the bunsen burner but the temperature remains steady at 100°C (Fig. 13.1). What is happening to the heat energy being supplied by the bunsen burner when the temperature remains at 100°C ?

Before the temperature reached 100°C the heat energy was raising the

fig 13.1 graph showing temperature against time for a beaker of heated ice/water



temperature of the water. When it reaches 100°C the heat energy is changing the state of the water; water is changing to steam and heat energy is needed to do this. The heat energy needed to change the state of a substance, that is to change it from solid to liquid or liquid to vapour, is known as the *latent heat*.

In order to change from the liquid state to the vapour state, the molecules must be given more energy so that they can break away from the attractive forces between them. The energy needed to do this is supplied by the heat energy. When the water is at 100°C , the further heat energy supplied to the water gives the molecules more kinetic energy, they move faster and can escape into the atmosphere.

The quantity of heat energy needed to change the state of 1 kg of a substance without change in temperature is known as the specific latent heat of the substance. It is measured in J/kg

Some typical values of specific latent heats are given in Table 13.1. The latent heat of fusion is the heat energy needed to change a solid into a liquid, and the latent heat of vaporisation is the heat energy needed to change a liquid into a vapour.

Table 13.1 *specific latent heats (J/kg)*

Fusion		Vaporisation	
Lead	21 000	Ether	370 000
Copper	181 000	Sulphur	1 510 000
Ice	336 000	Water	2 260 000

13.2 MEASURING SPECIFIC LATENT HEAT

The specific latent heat of vaporisation of water may be determined by boiling water in a beaker. A convenient way of boiling the water is to put an immersion heater in it. The energy supplied by the immersion heater may be determined by connecting the heater to the electrical supply via a joulemeter.

The experiment is illustrated in Fig. 13.2. The mass of water boiled off by a certain quantity of energy is determined using the balance. A typical set of readings is shown below.

Readings

Mass of water boiled away = 200 g

Heat supplied by immersion heater = $4.6 \times 10^5 \text{ J}$

Calculation

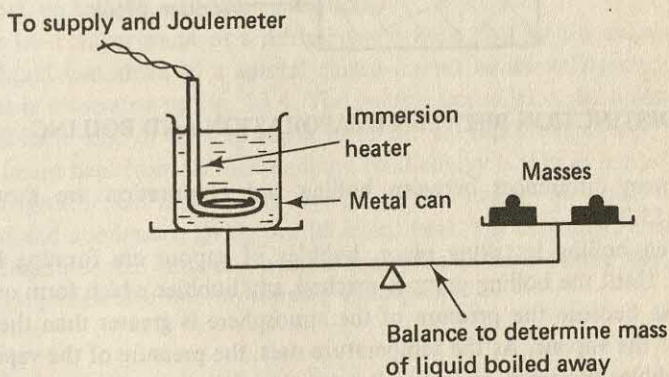
$$0.2 \text{ kg water are boiled away by } 4.6 \times 10^5 \text{ J}$$

$$1.0 \text{ kg water is boiled away by } \frac{4.6 \times 10^5 \text{ J}}{0.2}$$

$$= 2.3 \times 10^6 \text{ J}$$

$$\therefore \text{ Specific latent heat of vaporisation of water } = 2.3 \times 10^6 \text{ J/kg}$$

fig 13.2 *apparatus with which to measure the specific latent heat of vaporisation of water*

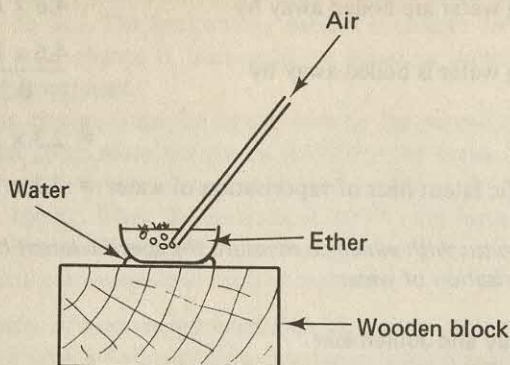


13.3 WHEN A LIQUID EVAPORATES, IT COOLS

In the experiment illustrated in Fig. 13.3, some ether is placed in a metal crucible and the metal crucible placed on a drop of water which is on a wooden block. When air is blown through the ether, the ether evaporates and the water freezes. The evaporating liquid has taken up its latent heat and this heat was supplied by the water. Heat energy is thus withdrawn from the water and the water freezes.

We may use the molecular theory of matter to explain the cooling (see Chapter 11). The average energy of molecules in a liquid is a measure of the temperature of the liquid. If the average energy of the molecules decreases, the temperature falls. When air is blown through the ether, this helps the faster-moving molecules to escape from the surface and the average kinetic energy of the remaining molecules is thus decreased. The temperature therefore falls.

If a drop of ether or petrol is placed on the hand, it immediately feels very cool. This is because the ether or petrol evaporates and the latent heat needed for it to evaporate is extracted from the hand.

fig 13.3 *the cooling caused by evaporating the ether can freeze the water*

13.4 DISTINCTION BETWEEN EVAPORATION AND BOILING

The main differences between boiling and evaporation are shown in Table 13.2.

When boiling is taking place, bubbles of vapour are forming in the liquid. Until the boiling point is reached, any bubbles which form quickly collapse because the pressure of the atmosphere is greater than the pressure of the vapour. As the temperature rises, the pressure of the vapour in the bubbles increases, and when it is equal to the atmospheric pressure the bubbles can grow. Boiling is then taking place.

Table 13.2 *differences between boiling and evaporation*

<i>Boiling</i>	<i>Evaporation</i>
Takes place at a definite temperature called the boiling point	Takes place at any temperature
Takes place in the body of the liquid. Bubbles can be seen forming throughout the liquid	Takes place only at the surface

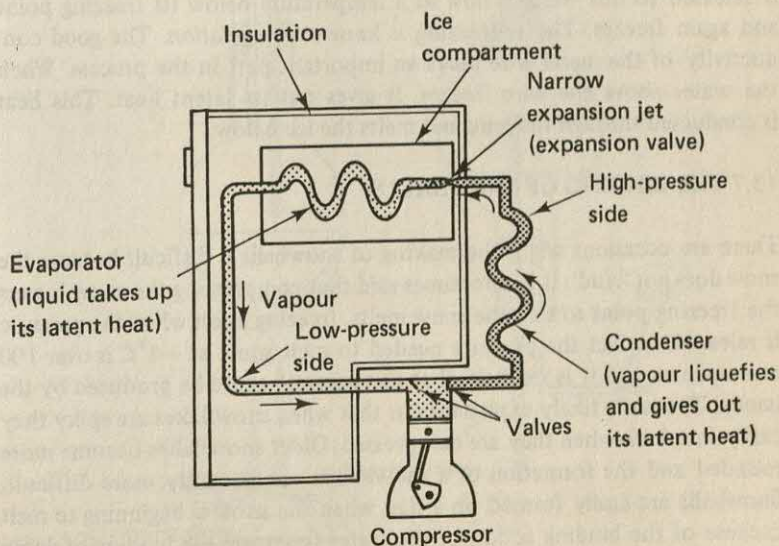
Up a mountain, where the atmospheric pressure is lower than at sea level of the Earth, the boiling point of a liquid is lower. The vapour pressure reaches the atmospheric pressure at a lower temperature and hence the liquid boils at a lower temperature. An experiment to illustrate this is described in Section 13.8.

13.5 THE BODY'S COOLING SYSTEM AND REFRIGERATORS

The evaporation of perspiration from the body helps to cool the body and keep it at a constant temperature. This evaporation prevents overheating of the body which, if it occurred, could cause irreversible damage and even death. Evaporation of water (or rather, the watery liquid secreted by the sweat glands) from the surface of the body is affected by the surrounding temperature, humidity (that is, the amount of water vapour in the atmosphere) and air currents. At an air temperature of 21°C a man can lose up to 25% of his heat energy loss by evaporation. (The great majority of his heat loss, up to 60%, will be by radiation.)

The ice compartment of a refrigerator is kept cool by the evaporation of a liquid contained in a special closed circuit in the refrigerator. The process is illustrated in Fig. 13.4. The compressor reduces the pressure on the left-hand side of the expansion valve and the liquid evaporates, taking up its latent heat from its surroundings. Heat energy is thus extracted from the refrigerator. On the opposite side of the compressor, the liquid is compressed and condenses, giving out its latent heat. The condenser is outside the refrigerator and thus this heat, having been extracted from the refrigerator, is passed into the surrounding atmosphere.

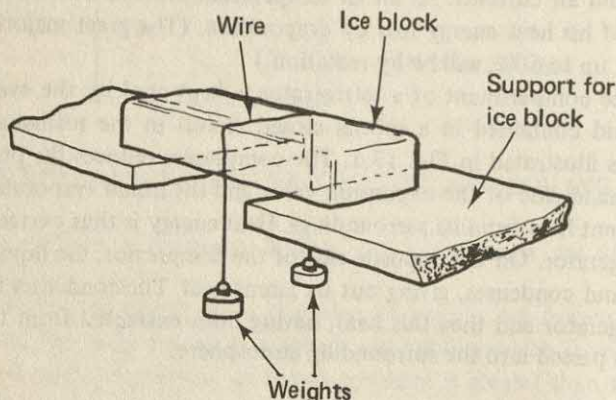
fig 13.4 *the principle of operation of the refrigerator. By evaporation and condensation round a cycle the refrigerant extracts heat energy from the insulated compartment and delivers it to the outside*



13.6 AN INCREASE IN PRESSURE LOWERS THE MELTING POINT

In the experiment illustrated in Fig. 13.5 a thin wire with heavy weights on each end is hung over a piece of ice. The wire exerts a large pressure on the ice and gradually passes through the block, yet the block still

fig 13.5 *the wire gradually cuts its way through the ice block, which re-freezes after the wire has passed*



remains a solid piece of ice. Why is this? The reason is that the large pressure on the ice lowers the melting point and the ice melts. Once the wire has passed through, the pressure on the water now above the wire is released so this water is now at a temperature below its freezing point and again freezes. The re-freezing is known as *regelation*. The good conductivity of the metal wire plays an important part in the process. When the water above the wire freezes, it gives out its latent heat. This heat is conducted through the wire and melts the ice below.

13.7 THE BINDING OF SNOWBALLS

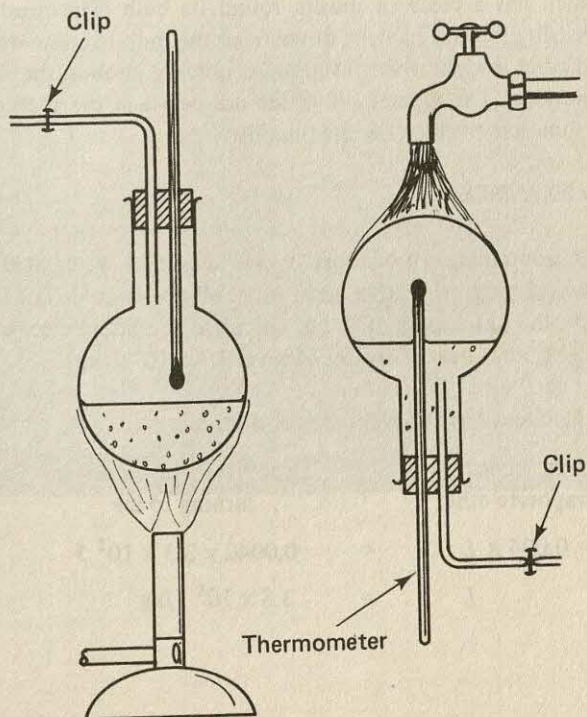
There are occasions when the making of snowballs is difficult because the snow does not 'bind'. It is sometimes said that compressing the snow lowers the freezing point so that the snow melts, freezing again when the pressure is released. In fact the pressure needed to melt snow at -1°C is over 100 atmospheres and it is unlikely that this pressure could be produced by the hands. The more likely explanation is that when snowflakes are spiky they easily interlock when they are compressed. Older snowflakes become more rounded and the formation of a snowball is consequently more difficult. Snowballs are easily formed on a day when the snow is beginning to melt because of the binding action of the water (compare the binding of damp and dry snow).

Skating is another example where melting and freezing are important. When an ice skate runs across an ice surface there is friction between the skate and the ice. This results in frictional heating and melting of a thin layer of ice just below the skates. There is therefore a lubricating layer of water for the skates to glide over.

13.8 A DECREASE IN PRESSURE LOWERS THE BOILING POINT

The experiment illustrated in Fig. 13.6 shows the decrease in boiling point as the pressure is lowered. The water in the flask is boiled and the clip closed to seal the flask. The flask is then inverted under a (cold) tap. When the tap is turned on, the temperature of the flask and contents decreases and the water begins to boil. In this way water may be made to boil at a temperature as low as 30°C . How does pouring cold water onto a flask cause the water in it to boil? The answer is that when the water was boiled initially, most of the air in the flask was driven out by the steam. When the clip was closed the flask was therefore full of water and steam. When cold

fig 13.6 *this experiment demonstrates that water can boil at a temperature considerably below 100°C*



water is poured onto the flask, the steam condenses and the pressure inside the flask therefore falls. At a lower pressure the water boils at a lower temperature; so it boils in the flask.

13.9 PRESSURE COOKERS

Pressure cookers make use of the fact that with an increase in pressure water boils at a higher temperature. In many pressure cookers, the pressure inside can have a maximum value of about 2 atmospheres and at this pressure the boiling point of water is about 120°C . At this temperature the chemical processes known as cooking take place considerably faster than they do at 100°C .

13.10 WATER VAPOUR IN THE ATMOSPHERE

When the amount of water vapour in the atmosphere is high, we say the *humidity* is high, and under these conditions we feel hot and uncomfortable. This is because perspiration from the body does not evaporate so readily as on a less humid day.

Humidity may be measured by an instrument known as an *hygrometer*. A common form of hygrometer consists of two mercury thermometers, one of which has a piece of muslin round its bulb. The muslin is kept moist by leaving the end hanging down from the bulb in some water. When the humidity is low the water evaporates quickly cooling the mercury in the thermometer. The greater the difference between the readings on the two thermometers the lower is the humidity.

WORKED EXAMPLE

A test-tube containing 4 g of ether is in a beaker of water at 0°C . Air at 0°C is blown through the ether until it has all evaporated. The ice formed on the test-tube has a mass of 4.2 g. Calculate the latent heat of evaporation of ether (latent heat of fusion of ice is $3.3 \times 10^5 \text{ J/kg}$).

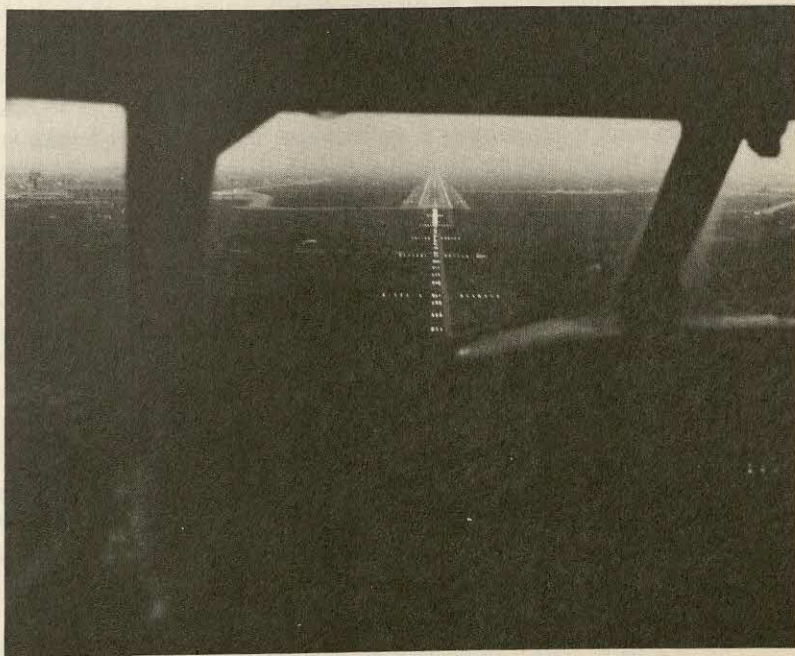
Let L be the latent heat of evaporation of ether.

$$\begin{array}{rcl}
 \text{Heat needed to} & = & \text{Heat given out by water} \\
 \text{evaporate ether} & & \text{turning to ice} \\
 0.004 \times L & = & 0.0042 \times 3.3 \times 10^5 \text{ J} \\
 L & = & 3.5 \times 10^5 \text{ J/kg}
 \end{array}$$

PART III

WAVE MOTION, LIGHT AND SOUND

Approach and runway lights as seen from the flight deck of an aircraft (courtesy British Aerospace)



WAVES

14.1 INTRODUCTION

A wave is a travelling disturbance which carries energy away from its source. Some waves, such as light waves and radio waves, can pass through an evacuated or empty space. In this way we receive energy from the Sun and other more distant stars.

Sound waves do not pass through a vacuum. Like some other waves, sound needs a medium or substance through which to travel; and it readily passes through many substances.

When waves pass through matter, the medium itself does not flow along in the direction of the waves. Water waves on a swimming pool, for example, do not cause the water to pile up at one edge of the pool; and a floating stick only bobs up and down as ripples flow past it on the water's surface. The stick does not move along with the waves.

14.2 DESCRIBING WAVES

(a) Frequency

When the term *frequency* is applied to waves, it refers to the rate at which they are produced by their source.

The frequency of a source is the number of waves it emits every second

The term is also used to refer to the number of waves reaching an observer every second. It is measured in hertz (Hz) or cycles per second. One particular radio transmitter for example, broadcasts by emitting 200 000 radio waves per second; so its frequency is 200 kHz. Radio receivers pick up the transmission, provided they are tuned to this frequency. The frequency of waves does not change as they travel away from their transmitter. It is an unchanging quality retained by the waves as they travel, even when they pass through different substances.

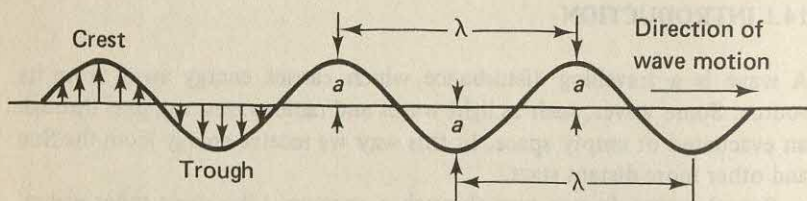
(b) Wavelength

Fig. 14.1 shows a cross-section of water waves made up of alternate crests and troughs flowing from left to right in the figure.

The wavelength, λ , of the waves is the distance between consecutive crests or consecutive troughs

Water waves can have wavelengths as small as a few millimetres or as large as a hundred metres or more. The wavelength of waves can change as

fig 14.1 *the amplitude a and wavelength λ of transverse water waves are shown. The vertical arrows show by how much the water's surface is displaced.*



they flow from one medium to another (Section 14.9(b)); but where the waves travel at a constant speed, their wavelength does not change.

(c) Amplitude

The term *amplitude* is used to describe the size of a wave. It is represented by the letter a in Fig. 14.1 where it measures the height of a crest or the depth of a trough from the average or undisturbed level of the water.

The amplitude of waves is their biggest displacement from the undisturbed level

Waves at sea have sometimes been described as four metres high, as their crests stood four metres above the level of their troughs. The amplitude of these waves would have been only two metres however. Can you see why?

14.3 TRANSVERSE WAVES

The disturbances in transverse waves take the form of displacements perpendicular to the direction in which the waves travel

In Fig. 14.1, for example, the surface of the water is displaced vertically as shown by the arrows; but the crests and troughs travel horizontally, that is, in a perpendicular direction. These are transverse waves; and so are

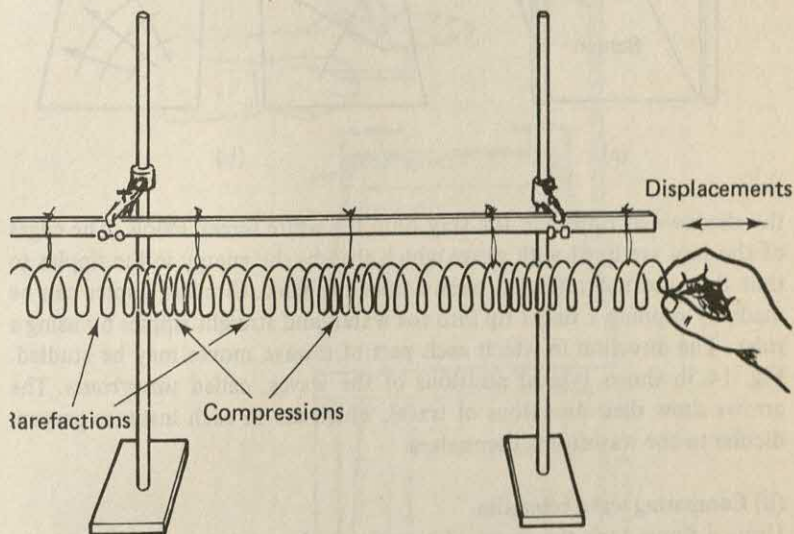
most of the waves we come across. Radio waves and light are typical transverse waves; so also are the sideways ripples you can make on a stretched rope.

14.4 LONGITUDINAL WAVES

In longitudinal waves the disturbances are displacements backwards and forwards along the direction in which the waves travel

Sound waves and compression waves are longitudinal; Fig. 14.2 illustrates this type of wave in a 'slinky' supported by strings at various

fig 14.2 *compression waves in a slinky are caused by longitudinal displacements of one end of the spring*



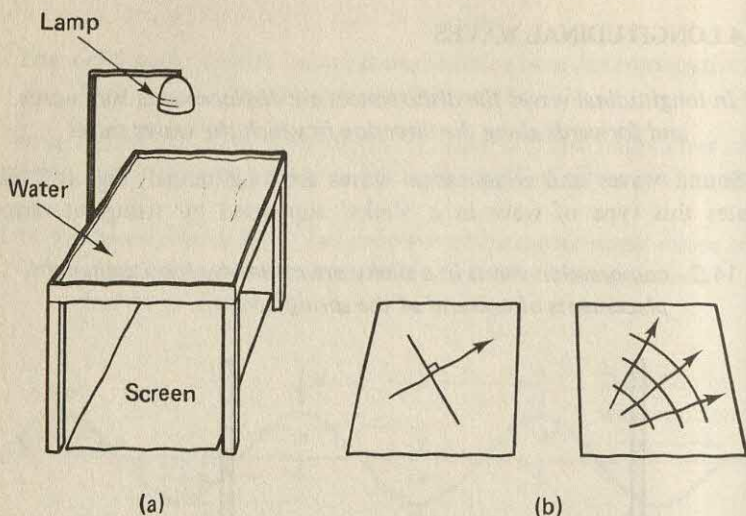
intervals along its length. If one end of the slinky is displaced sharply backwards and forwards along its length, a compression and then a decompression or rarefaction can be seen moving along the spring.

14.5 OBSERVING AND MEASURING TRAVELLING WAVES

(a) Seeing how the waves move

The motion of water waves may be studied by looking at ripples in a ripple tank. This is a shallow tray of water supported on legs above a horizontal screen of white paper. The bottom of the tray is clear glass, and above it there is a small lamp as shown in Fig. 14.3a. In a dimly lit room the lamp projects

fig 14.3 (a) a ripple tank. (b) Waves move in directions perpendicular to wavefronts



the shadows of ripples in the tray onto the white screen below. The edges of the tray are lined with gauze which absorbs the energy in the ripples so that they are not reflected back across the tank. Circular ripples can be made by dipping a finger-tip into the water, and straight ripples by using a ruler. The direction in which each part of a wave moves may be studied. Fig. 14.3b shows typical positions of the waves, called *wavefronts*. The arrows show their directions of travel, which are in each instance perpendicular to the wavefronts themselves.

(b) Comparing wave velocities

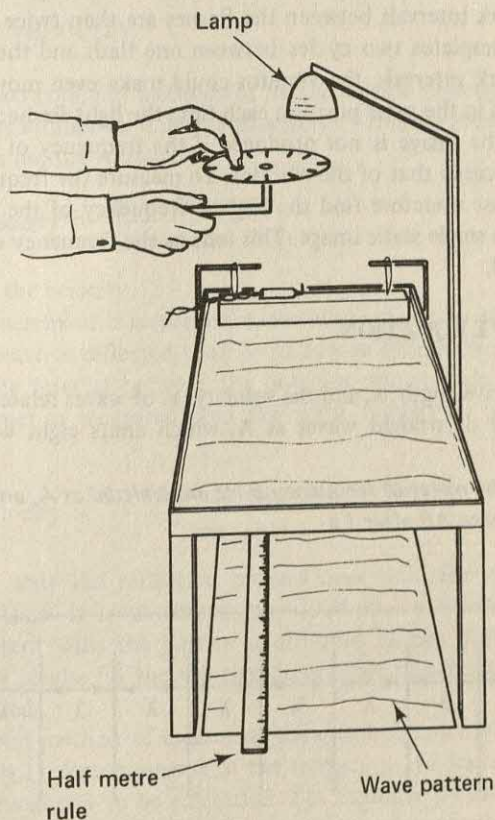
Using a ripple tank the speed of a single ripple can be compared with the speed of a group of several ripples close together. The back ripples in the group neither catch up nor fall behind the front ones, as all the ripples move with the same speed.

The tank may be tilted by putting books under the legs at one end. The water is then deeper at one end of the tank. In deep water the waves are seen to move faster than in shallow water. As the waves move from the deep end to the shallow end, their speed decreases and their wavelength gets shorter. Light waves behave in a similar way and are slowed down when they pass from air into glass, but they regain their original speed when they emerge from glass back into air. It is the medium which determines the speed at which the waves travel (Section 14.9(b)).

(c) Measuring the wavelength of waves

In a level ripple tank equipped with a regular vibrator, the ripples are uniformly spaced. Their apparent wavelength is measured using a *stroboscope* and a rule. The stroboscope is a circular disc with regularly spaced radial slots cut in its edge. It is held by a handle and the disc is rotated beneath the lamp of the ripple tank, as illustrated in Fig. 14.4. This regularly

fig 14.4 a stationary pattern of wave shadows is produced on the screen with the aid of a hand-held stroboscope



interrupts the view of the shadows on the screen below. If the disc's rate of rotation is adjusted correctly, the intervals in which the light is blocked off are just enough for one ripple to move into the position previously taken up by its neighbour. In this way the shadows of the ripples on the screen appear to stand still, and their wavelength can be measured using the half-metre rule.

(d) Measuring the frequency of waves

Waves generated by an electrical vibrator have the same frequency as their source. A flashing light stroboscope may be used for measuring this frequency. It has a light which can be flashed many times a second at an adjustable and known rate. In this light, ordinary movements appear jerky: it is used here to make the vibrating source of waves appear at rest. If the operating frequencies of the vibrator and stroboscope coincide, the vibrator completes just one cycle in each interval between the flashes of light, so the vibrator is repeatedly seen in the same position.

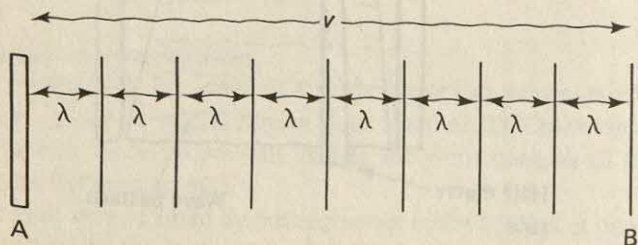
The same stationary image is seen if the stroboscope frequency is halved. The dark intervals between the flashes are then twice as long, and the vibrator completes two cycles between one flash and the next. With even longer dark intervals, the vibrator could make even more vibrations and still be seen in the same position each time the light flashed.

A single static image is not produced if the frequency of the flashing stroboscope exceeds that of the vibrator. To measure the frequency of the vibrator we must therefore find the highest frequency of the stroboscope that results in a single static image. This tells us the frequency of the waves being generated.

14.6 THE WAVE EQUATION

How are the wavelength, λ , and the velocity, v , of waves related? Fig. 14.5 shows a source of straight waves at A, which emits eight waves in one

fig 14.5 *straight waves of frequency 8 Hz are emitted at A, and they fill the space AB after 1 s*



second. Their frequency is therefore 8 Hz. Their forward velocity is v m/s, so the first wave has reached the position B after 1 s, where $AB = v$ m. All eight wavelengths are contained in the distance AB, so the wavelength is $v/8$ m. If the frequency had been 16 Hz, then all sixteen wavelengths would have been contained in v m after 1 s, and their wavelength would

have been $v/16$ m. In general where the frequency of waves is f Hz, their wavelength can be written as

$$\lambda = v/f$$

or

$$v = f\lambda \quad (14.1)$$

This is called the *wave equation*. It applies to all types of waves.

14.7 WAVE VELOCITY

There are two common methods of measuring the velocity of waves. The first requires a timer to time the passage of a single wave over a measured distance. We use the equation

$$\text{Velocity} = \frac{\text{Distance gone}}{\text{Time taken}} \quad (14.2)$$

to calculate the velocity.

The measurement is sometimes easier in practice if an echo technique is used. The wave is reflected back to its source from a barrier at a known distance. The interval between the outgoing and reflected waves as they pass the timer is measured, and the wave velocity is found from the equation

$$\text{Velocity} = \frac{2 \times \text{Distance from timer to barrier}}{\text{Time taken}} \quad (14.3)$$

We assume that the reflection process does not take up a significant amount of time. It is an assumption which can be checked by repeating the experiment with the barrier at different known distances from the timer. If the results for the velocity do not vary significantly, the assumption is justified.

The second method of measuring the speed of waves relies on the wave equation (14.1). Measurements of the frequency and wavelength of waves enable their velocity to be calculated. For example, we can find the speed of radio waves in air from the knowledge that the transmission of frequency 200 kHz has a wavelength of 1500 m. Since

$$\text{Velocity} = \text{Frequency} \times \text{Wavelength}$$

we have

$$\begin{aligned} v &= 200\,000 \times 1500 \text{ m/s} \\ &= 2 \times 1.5 \times 10^8 \text{ m/s} \\ &= 3 \times 10^8 \text{ m/s} \end{aligned}$$

No waves have been found to carry energy faster than this. This speed is the speed of light and of radio waves in free space. Sound waves travel much more slowly than this in air, as we know from the time interval between the lightning we see and the thunder we hear.

14.8 ENERGY IN WAVES

Waves carry energy, and a source therefore uses up energy when it creates a wave. Big waves carry more energy than small ones, and experiments have shown that in a given length of wavefront the energy carried is proportional to the (wave amplitude)².

As a circular water ripple spreads out, its amplitude decreases. This is true of all waves spreading out from their sources. The energy the waves carry is distributed along an ever-increasing wavefront, and eventually it may become so spread out that the waves can no longer be detected.

14.9 PROPERTIES OF TRAVELLING WAVES

All known wave motions share four common properties. They can be reflected, refracted and diffracted, and can exhibit interference.

(a) Reflection

Fig. 14.6a shows how plane (straight) waves rebound from a straight barrier. The waves retain their speed and wavelength, as they remain in the same medium. Their directions of motion before and after reflection make equal angles with the 'normal', which is drawn at 90° to the barrier. The waves themselves lie perpendicular to their directions of motion.

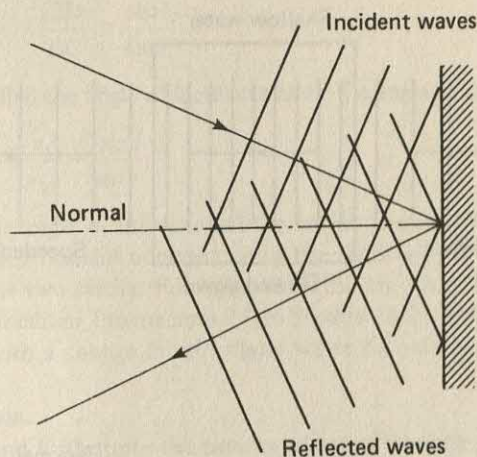
Fig. 14.6b shows circular waves rebounding from a straight barrier. Again their wavelength and speed are unchanged: the curvature of the waves is altered, however, so that the reflected waves are centred at S', which is the 'image position' of the source S behind the barrier. Reflection is the subject of Chapter 15.

(b) Refraction

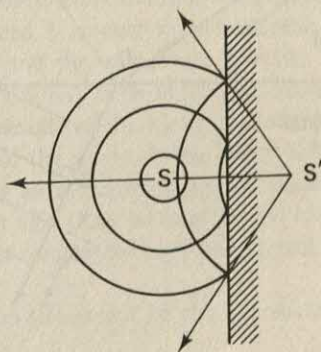
On passing from one medium to another, waves undergo changes of speed and wavelength; but their frequency remains unchanged. Fig. 14.7a shows a special instance of this when straight ripples enter a region of shallower water. Changes of speed and wavelength occur at the boundaries between the shallow and the deeper regions, but the waves continue to move in the same direction as before. The wavelength of the waves is greatest where they travel fastest.

Fig. 14.7b illustrates the effect when straight ripples meet the boundary between the shallow and deeper water obliquely. In addition to the

fig 14.6 (a) the directions taken by incident and reflected waves make equal angles with the normal to the reflecting surface. (b) After circular waves from S are reflected at a straight barrier, they travel as if they had come from S'



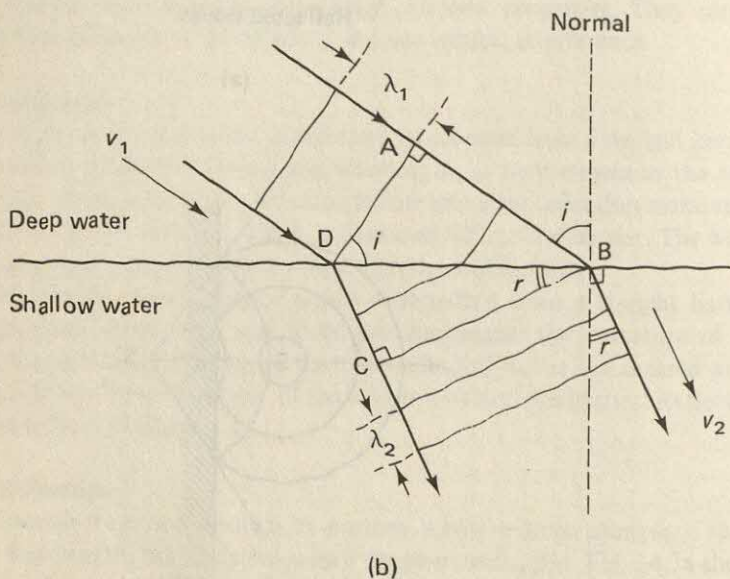
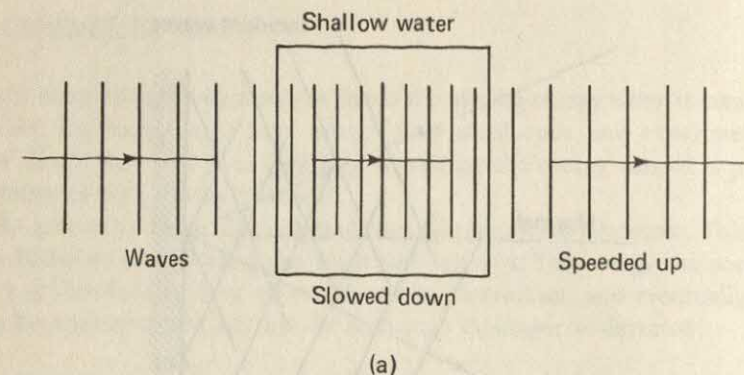
(a)



(b)

previous effects we can see the marked change in the direction of travel of the waves. The wavefront AD becomes the wavefront BC in the shallower water. The distances AB and DC are travelled in the two media in equal times, and their ratio is the ratio of the two wave speeds $v_1:v_2$. As AB is greater than DC , we can tell that v_1 is greater than v_2 .

fig 14.7 (a) waves travelling along the normal to a boundary are not deviated when they undergo changes of speed and wavelength.
 (b) Waves crossing a boundary obliquely are refracted and travel in a new direction



The wave equation $v = f\lambda$ applies to the waves each side of the boundary BD where they have the same frequency; it follows that their wavelengths are in the same ratio as their velocities:

$$\frac{\lambda_1}{\lambda_2} = \frac{v_1}{v_2} = \frac{AB}{DC} \quad (14.4)$$

This ratio could also be written

$$\frac{AB}{DC} \times \frac{BD}{BD} \quad \text{or} \quad \frac{AB}{BD} \times \frac{BD}{DC}$$

In the right-angled triangles ABD and CDB in Fig. 14.7b, the ratio

$$\frac{AB}{BD} \times \frac{BD}{DC} = \frac{\sin i}{\sin r}$$

where i is called the angle of incidence and r the angle of refraction:

$$\therefore \frac{\lambda_1}{\lambda_2} = \frac{v_1}{v_2} = \frac{\sin i}{\sin r} \quad (14.5)$$

Whatever the angle of incidence of the waves crossing this boundary, the ratio $\sin i / \sin r$ remains constant, since the ratio v_1/v_2 is a constant determined by the two media. For light waves the ratio is called the refractive index from medium 1 to medium 2 (see Section 16.2). Refraction is always associated with a change in velocity of waves on entering a new medium.

(c) Diffraction

Fig. 14.8a and b illustrate the passage of waves through gaps in a barrier. Waves in a ripple tank passing through the centre of a gap continue straight on; but there is a tendency for the end of each wavefront to deviate into the geometrical shadow area behind the barrier. The waves remain in the same medium, so there is no change in their speed or wavelength. In Fig. 14.8a where the gap is much wider than the wavelength of the waves, the effect is not very pronounced. It is much more noticeable in Fig. 14.8b, where the width of the gap and the wavelength of waves are similar. In general, we can say that there is very little diffraction when apertures are wide compared with the wavelength of the waves. A similar situation arises if an object blocks the path of the waves. If the object is large compared with the wavelength, there is an effective region of 'shadow', but if the object is small the diffraction effect can be so great that after a few wavelengths the wavefront appears completely undisturbed, just as if there had been no object in its path.

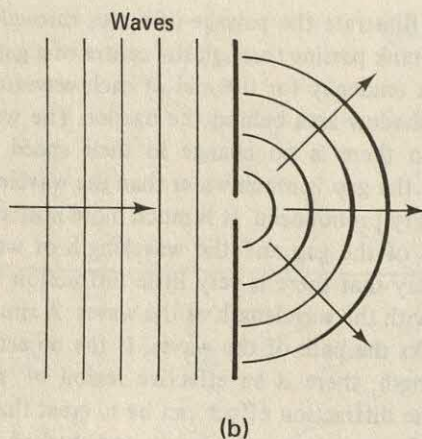
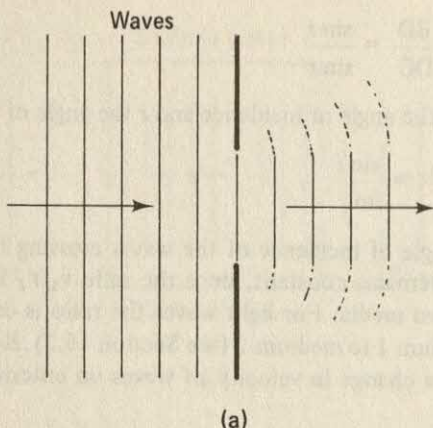
All waves exhibit diffraction effects and we now discuss three examples.

(i) Visible light

In air, visible light has wavelengths close to 5×10^{-7} m. The wavelength varies with colour, however, that of red light being about 1.7 times the wavelength of blue light.

To observe the diffraction of light, we view a distant filament lamp through a narrow scratch on a blackened microscope slide. The scratch

fig 14.8 (a) waves passing through a gap much wider than their wavelength show slight diffraction effects. (b) Waves passing through a gap comparable in width with their wavelength show marked diffraction effects



which lets through the light is set parallel to the long bright filament of the bulb (Fig. 14.9). Fig. 14.10 illustrates the patterns of light and dark bands which the observer sees when first a red filter, and then a blue filter is placed in the path of the light. These are called 'single-slit diffraction patterns'. Each is comprised of several light and dark bands with a broad central band of colour twice as wide as its neighbours on either side. Apart from their colour, the chief difference between the patterns is in the widths of the bands. The red bands of colour are about 1.7 times as wide as the corresponding bands in the blue pattern. This is entirely due

fig 14.9 *a tall filament is viewed through a narrow scratch on a darkened slide. Light from the filament is diffracted on passing through the slit producing a single-slit diffraction pattern*

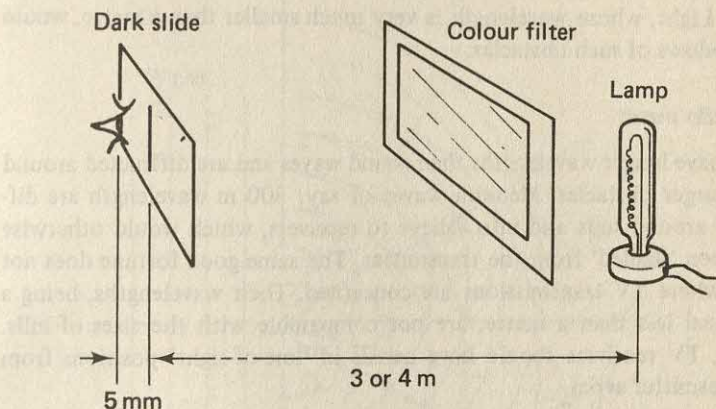
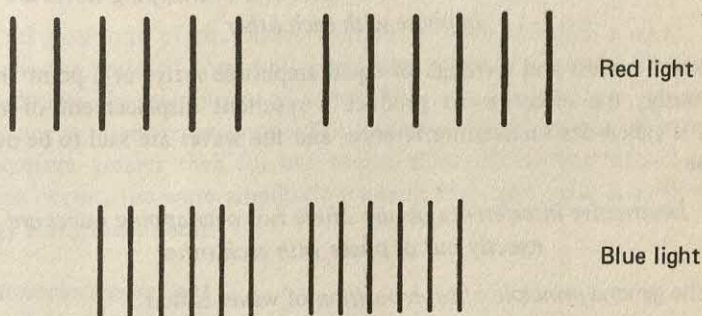


fig 14.10 *single-slit diffraction patterns in red and blue light each have a broad central band of light with narrower bands on each side. If the wavelength of red light is 1.7 times that of blue light, then the bands of red light are 1.7 times as wide as the corresponding bands in the blue diffraction pattern*



to their wavelength ratio (1.7); a comparison of unknown wavelengths can be made by this method.

Closer study shows that the widths of the bands of light in a diffraction pattern depend on both the wavelength of the light and the width of the scratch on the slide. If we increase the wavelength for a given width of scratch, the pattern widens; if we use a wider scratch and the same wavelength, the pattern narrows.

(ii) Sound waves

These are diffracted around obstacles and are clearly heard on the opposite side to the source (e.g. round corners of buildings). The effect is heard best if obstacles are about a metre across, as this is similar to the wavelength of sound. Light, whose wavelength is very much smaller than a metre, would cast shadows of such obstacles.

(iii) Radio waves

These have longer wavelengths than sound waves and are diffracted around much larger obstacles. Medium waves of say, 300 m wavelength are diffracted around hills and into valleys to receivers, which would otherwise have been 'shaded' from the transmitter. The same good fortune does not apply where TV transmissions are concerned. Their wavelengths, being a good deal less than a metre, are not comparable with the sizes of hills. Ideally, TV receivers should have aerials in 'line-of-sight' positions from the transmitter aerial.

(d) Interference

Interference occurs where two or more similar sets of waves are superimposed on each other. The result of two wave crests arriving simultaneously at a point is a larger crest; and two troughs arriving together result in a larger trough. This is called constructive interference, and the waves are said to be 'in phase'.

Constructive interference occurs where two overlapping waves are in phase with each other

Where a crest and a trough of equal amplitude arrive at a point simultaneously, the effect is to produce a resultant displacement of zero. This is called destructive interference, and the waves are said to be out of phase.

Destructive interference occurs where two overlapping waves are exactly out of phase with each other

The general principle of superposition of waves is that

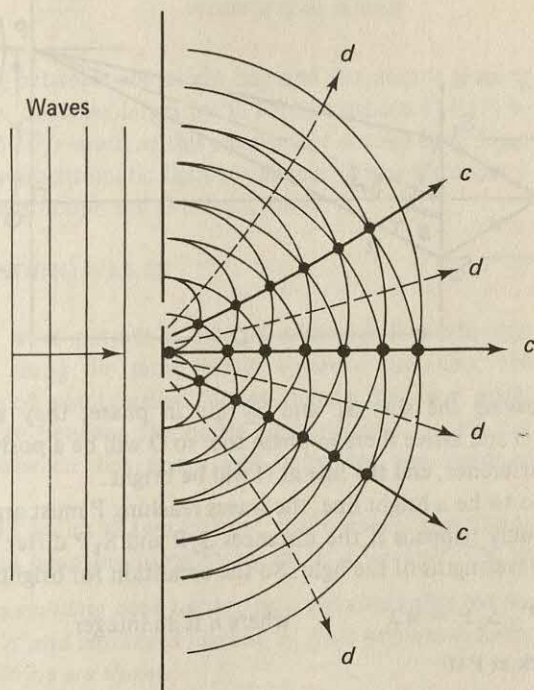
The total displacement at a point at any instant is the sum of the individual wave displacements arriving at the point at that time

We now consider two particular cases of interference.

(i) Water ripples

These create an interference pattern in a ripple tank equipped with two small round dippers and an electrical vibrator. Fig. 14.11 illustrates instan-

fig 14.11 an interference pattern of waves emerging in phase from two identical slit sources. Paths marked 'c' indicate constructive interference and those marked 'd' indicate destructive interference

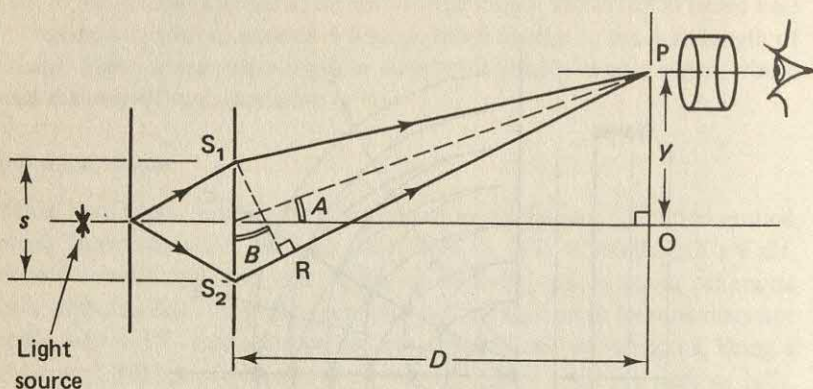


taneous positions of the waves emitted by two identical sources, and it shows the paths of constructive and destructive interference. The waves have double the amplitude of waves from one source alone, wherever constructive interference occurs. The rate of energy flow along these paths is therefore greater than for one source alone. Where destructive interference occurs, the wave amplitude tends to zero, and there is no flow of energy along these paths.

(ii) *Monochromatic light*

This is light of one colour, and so it is all of one wavelength, λ , in air. In the experiment referred to as 'Young's slits', monochromatic light passes through a single slit and then through two narrow parallel slits no more than half a millimetre apart. Beyond the double slits the light emerges in two sets of waves which overlap and cause interference. In Fig. 14.12 the interference pattern is examined through an eyepiece focused in the plane OP, where a series of light and dark parallel lines or 'fringes' can be seen. If

fig 14.12 *Young's slits apparatus, which is not drawn to scale. Light and dark fringes are observed in the plane OP, resulting from the interference of light from the slits S_1 and S_2*



the waves leaving the slits S_1 and S_2 are in phase, they travel equal distances to O and arrive there in phase too; so O will be a position of constructive interference, and the line at O will be bright.

If P is also to be a bright line, the waves reaching P must arrive there in phase. This only happens if the distances S_2P and S_1P differ by a whole number of wavelengths of the light. So the condition for brightness at P is:

$$S_2P - S_1P = n\lambda \quad \text{where } n \text{ is an integer} \quad (14.6)$$

It will be dark at P if:

$$S_2P - S_1P = (n + \frac{1}{2})\lambda \quad (14.7)$$

Fig. 14.12 is not drawn to scale for the sake of clarity. In reality the distances s , y , and the wavelength λ , are all very much less than the distance D ; and the angles A and B are small and approximately equal angles. In these conditions

$$S_2P - S_1P \approx S_2R \quad (14.8)$$

Using these approximations, we can write the condition for P to be bright as:

$$S_2R = n\lambda \quad (14.9)$$

As the angles A and B are small, and approximately equal

$$\tan A \approx \sin B \quad (14.10)$$

(the tangent and the sine of an angle are equal, for small angles.) Substituting, this becomes:

$$\frac{y}{D} \approx \frac{n\lambda}{s} \quad \text{if P is bright} \quad (14.11)$$

So the condition for P to be bright is:

$$y = \frac{n\lambda D}{s} \quad \text{where } n \text{ is an integer} \quad (14.12)$$

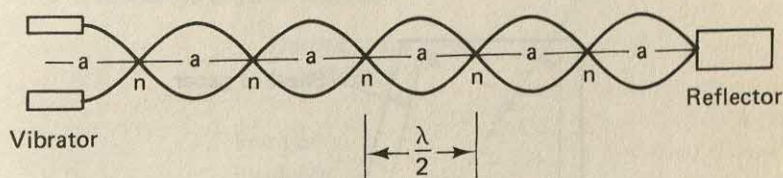
The spacing between one bright line and the next is given by the increase in y when n becomes larger by 1. From equation (14.12) we can see that the lines are $\lambda D/s$ apart, so this experiment enables us to measure the wavelength of monochromatic light by measuring the distances y and s using a travelling microscope and D using a ruler.

14.10 STANDING WAVES

A standing wave pattern is created whenever identical sets of travelling waves pass along the same line in opposite directions. The pattern has equally spaced points called 'nodes', where the wave amplitude remains zero. Midway between the nodes are 'antinodes', where the amplitude is greatest. Between the nodes and antinodes the amplitudes gradually increase.

Fig. 14.13 indicates two extreme positions of a string on which a standing wave has been produced by means of a vibrator. The waves reaching

fig 14.13 *a standing wave pattern on a vibrated string has nodes marked 'n' and antinodes marked 'a'. Two extreme positions of the string are shown*



the far end are reflected back along the string. The nodes in the figure are marked 'n', and the antinodes 'a'. The nodes are spaced at half wavelength intervals.

In the standing wave pattern the displacements between adjacent nodes are all in phase, but their amplitudes vary. The energy of this type of wave is contained mostly at the antinodes where the amplitude is greatest, and there is none at the nodes.

Standing waves are important in many musical instruments, as they give rise to the travelling sound waves which we hear. More about this can be found in Chapter 20.

REFLECTIONS AND SHADOWS

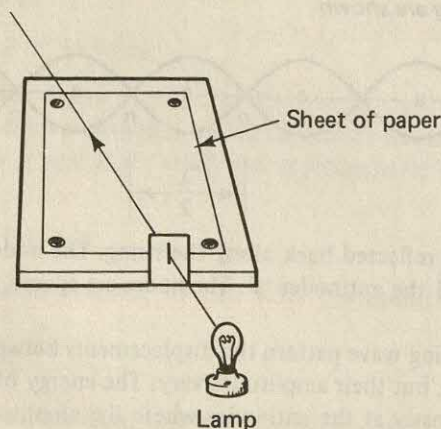
15.1 RAYS OF LIGHT

Light energy travels outwards from its source, and a line showing its direction of travel is called a *ray*. When light from a bright source passes through a narrow slit in front of it, the light is easily seen where it falls on a surface, for example a sheet of paper (Fig. 15.1). The bright line across the paper from the slit marks the path of the ray, which is a straight line.

Light travels in straight lines

In diagrams, rays are represented by unbroken lines each bearing an arrow pointing along the path travelled by the light.

fig 15.1 *the ray's path is a straight line*



15.2 LAWS OF REGULAR REFLECTION

Regular reflection occurs at surfaces so smooth that the path of a reflected ray is even more predictable than the path of a billiard ball rebounding

from a cushion at the side of the table. Glass and polished metal surfaces are normally good regular reflectors of light; you may be able to think of others.

With the apparatus illustrated in Fig. 15.2, rays can be plotted before and after they meet a plane mirror. The *incident ray* meets the mirror at the *point of incidence*, and the *reflected ray* marks the path of light away from the mirror after reflection. Fig. 15.3 shows the rays seen from above. The reflected ray lies in the plane of the paper, being neither directed up

fig 15.2 *apparatus for plotting the paths of incident and reflected rays near a plane mirror*

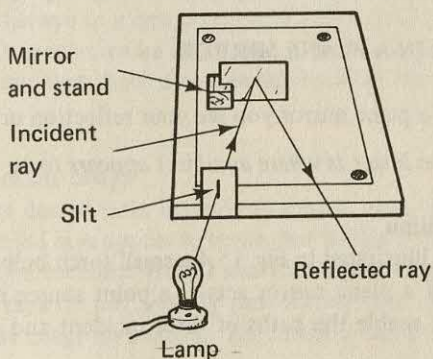
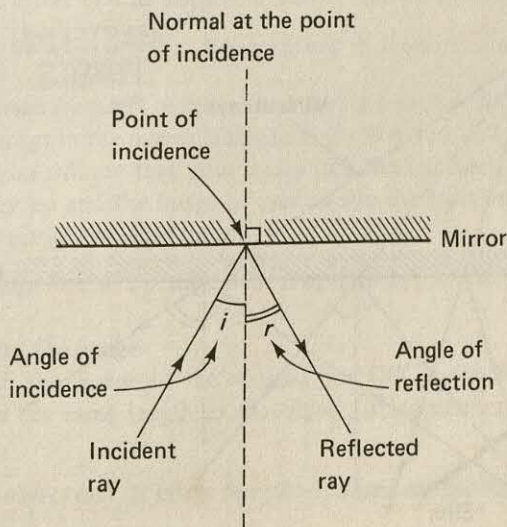


fig 15.3 *where a typical incident ray is reflected the angle of incidence equals the angle of reflection*



out of the page nor down into it. The normal to the surface at the point of incidence is a line perpendicular to the mirror. It lies between the two rays. These properties are stated in the *first law of reflection*:

The reflected ray lies in the plane formed by the incident ray and the normal to the surface at the point of incidence. The two rays lie on opposite sides of the normal

In Fig. 15.3 the *angle of incidence* i between the incident ray and the normal is found to equal the *angle of reflection* r between the reflected ray and the normal. The *second law of reflection* is:

The angle of incidence is equal to the angle of reflection

15.3 THE IMAGE IN A PLANE MIRROR

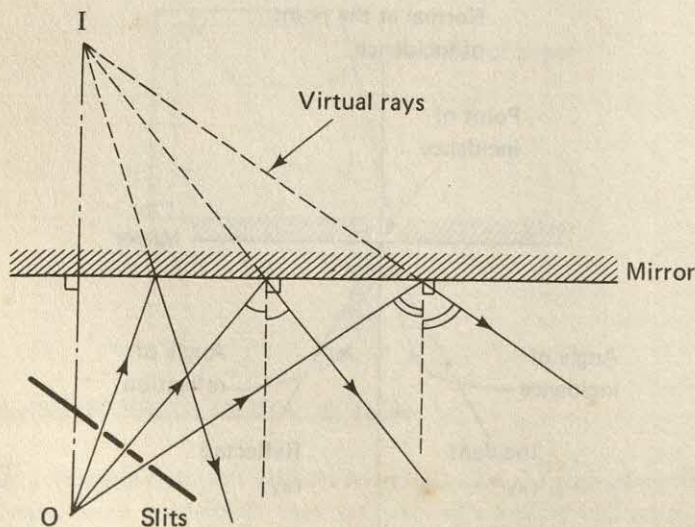
When you look at a plane mirror you see your reflection or *image*.

An image is where an object appears to be

(a) The image position

In the experiment illustrated in Fig. 15.4 a small torch bulb O placed about 10 cm in front of a plane mirror acts as a point source of light. Narrow slits in a flat card enable the paths of three incident and reflected rays to

fig 15.4 *the position of the image I is found by drawing the virtual rays behind the mirror to the point where they meet*



be plotted. It is found that the equal angles property is obeyed at each point where the rays are reflected. Dotted construction lines are drawn extending the reflected rays behind the mirror to the point where they meet. They appear to come from I, which is the image position of O. The line OI joining the point object to its image is a normal to the mirror's surface, and the mirror line bisects OI.

The image is as far behind a plane mirror as the object is in front of the mirror

The results can be tested by standing a tall vertical pin at O, and a second tall pin which can be seen above the mirror, at I. An observer sees the image of the object pin in line with the top half of the pin at I; and when the observer moves his eye to a new position, the alignment remains unchanged. The image of O appears to be at I, when it is viewed from many different directions. We say that there is *no parallax* between the object pin and the image pin.

(b) The nature of the image

In Fig. 15.4 the dotted lines behind the mirror mark the directions from which the reflected rays appear to come, but the light does not come from I. It only appears to come from I and the dotted lines are called *virtual rays*. Since the rays do not actually pass through the image, it is not possible to form this image on a screen. The image is called a *virtual image*.

Images behind a plane mirror are virtual images

(c) The appearance of the image

Look at the cover of this book in a mirror, and you will see that the title appears as

MISTARTAM
2012YH9

When viewed in a mirror everything on the

cover is reversed from left to right. The image is said to be *laterally inverted*. Your own image in the mirror is also laterally inverted, and if you lift your right hand you will see that your image lifts the left hand. Your image is the right way up and the image of print is also the right way up. They are described as *upright* or *erect* images.

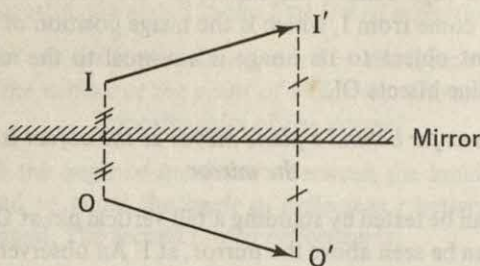
The image behind a plane mirror is upright but laterally inverted

(d) The size of the image

In Fig. 15.5 the image of the straight line OO' is the straight line II'; the image is the same length as the object, being neither magnified nor diminished.

The object and its image in a plane mirror are the same size

fig 15.5 *the image of the line OO' in the mirror is $I'I'$. The object and its image are the same length*



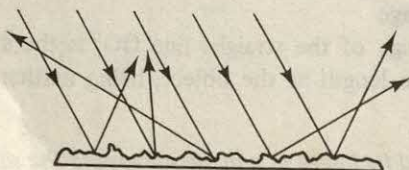
15.4 USES OF PLANE MIRRORS

Plane mirrors are commonly used in bedrooms and bathrooms and as hand mirrors. They are also used in scientific instruments which have pointers that move over a scale. In order to read the instrument correctly the eye must be directly over the pointer. Hold a pencil about 5 cm above a ruler in line with your eye. Then as you move your eye from side to side, the scale reading behind the pencil will appear to change, causing a reading error known as the *parallax error*. If an instrument is fitted with a plane mirror, the scale reading should only be taken when the image of the pointer in the mirror is hidden behind the pointer. This ensures that the observer's eye is correctly placed and an error due to parallax is avoided.

15.5 DIFFUSE REFLECTION

When light rays are reflected at a rough white surface, the reflection is no longer regular. In Fig. 15.1 the ray which falls on white paper can be seen by observers on all sides. The paper's surface receives the ray, and scatters its light making reflected rays in all directions. Fig. 15.6 shows how a rough or granular surface does this to a parallel-sided incident beam of light. This sort of reflection is called *diffuse reflection*.

fig 15.6 *an illustration of diffuse reflection where a parallel-sided incident beam is reflected at an irregular surface*

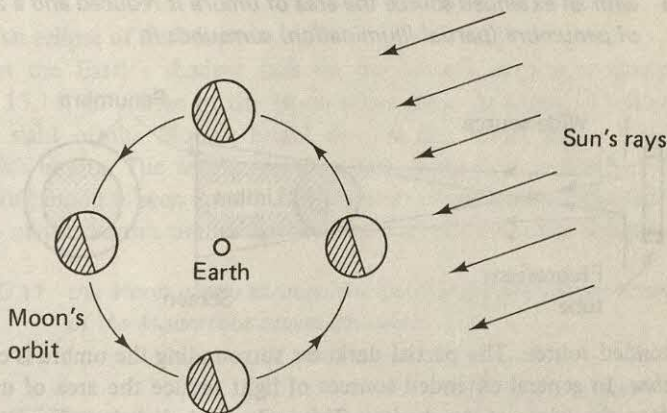


Diffuse reflection occurs at surfaces which scatter the light they receive in all directions

A cinema screen is a diffuse reflector. It receives light from a film projector, and scatters it so that people sitting anywhere in the cinema can see the picture focused on the screen. A painted wall with a matt surface also has this property, as tiny particles in the paint diffusely reflect the light.

The Moon's surface is a diffuse reflector of sunlight. At any instant, one-half of its surface is in direct sunlight, and the other half is in the shadow. The Moon's orbit takes it around the Earth in a lunar month, which is a period of about 28 days. The Moon's appearance varies according to how much of its illuminated surface is turned towards Earth. If you were at the centre of the Moon's orbit illustrated in Fig. 15.7, and you looked towards the Moon in each position shown in that figure, you would see one new Moon and one full Moon during the 28 day orbit.

fig 15.7 *the Moon's appearance from Earth varies during its 28 day orbit (not drawn to scale)*



15.6 SHADOWS

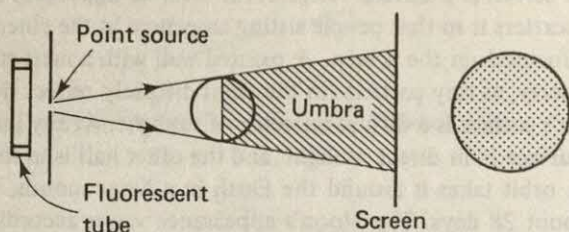
(a) Shadows cast by point sources of light

A point source of light casts a sharp shadow of an object onto a screen as in Fig. 15.8. The object in the figure is spherical, and its shadow appears on the screen as a dark circle on a light background. Shadow receiving no light from the source is called *umbra*.

(b) Shadows cast by extended sources of light

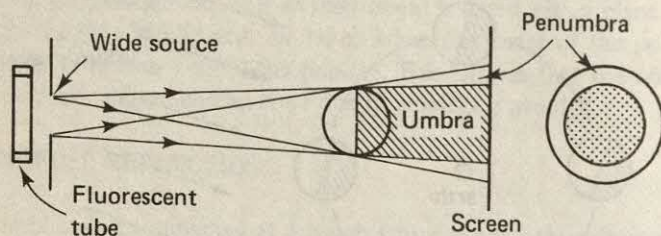
An extended source of light is a source which emits light in all directions from many different points. A frosted glass window, for example, and a

fig 15.8 *a circular sharp shadow of a sphere is cast by a point source of light*



strip lighting tube behave as extended light sources. Each point of an extended source casts a shadow behind the object; but the many shadows do not exactly coincide. The result is a shadow like that in Fig. 15.9. In the zone surrounding the umbra there is an area only partly lit by the source. It is partly lit as it receives light from only a fraction of the area of

fig 15.9 *with an extended source the area of umbra is reduced and a zone of penumbra (partial illumination) surrounds it*



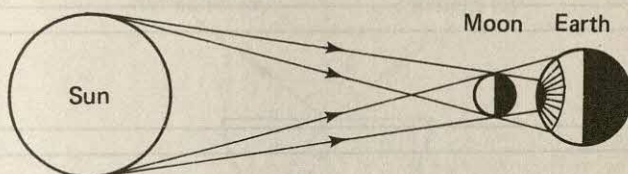
the extended source. The partial darkness surrounding the umbra is called *penumbra*. In general extended sources of light reduce the area of umbra and soften the edges of the shadow. This makes strip lighting very suitable for use in factories and workshops.

15.7 ECLIPSES

(a) An eclipse of the Sun

The Sun is an extended source of light casting both umbra and penumbra behind the Moon. These shadows seldom fall on the Earth's surface as the Sun, the Moon and the Earth rarely lie in a straight line; but when it does happen, the Sun is partly or totally lost from view as the Moon's umbra and penumbra pass across the Earth's surface (Fig. 15.10). The event, which occurs only at a time of new moon is called an eclipse of the Sun. Darkness falls on the Earth's surface where it would otherwise be daylight, and at

fig 15.10 *the Moon's umbra and penumbra pass across the Earth's surface during an eclipse of the Sun (not drawn to scale)*



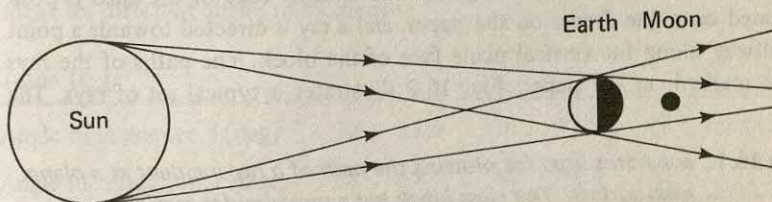
these places the Moon can be seen only as a dark circle covering all or part of the Sun. A total eclipse occurs where the Sun is totally lost from view.

Though total solar eclipses are rare at any one place, they have been the subject of man's interest and curiosity from ancient times. It is thought that the builders of Stonehenge knew enough to predict when a solar eclipse would be visible from there. Today's astronomers plan ahead, looking to the time of the next solar eclipse in order to be at the most suitable place on the Earth's surface to photograph the event.

(b) An eclipse of the Moon

When the Earth's shadow falls on the Moon's surface as illustrated in Fig. 15.11, an eclipse of the Moon takes place. At a time of full moon we lose sight of the Moon's bright disc, as the Moon passes through the Earth's umbra. The whole event can take more than an hour, after which the full moon is seen again in the night sky. During the eclipse the circular edge of the Earth's umbra can be seen as it falls across the Moon's surface.

fig 15.11 *the Moon passes through the Earth's shadow during an eclipse of the Moon (not drawn to scale)*



REFRACTION

16.1 INTRODUCTION

A swimming pool viewed from a spring board vertically above its surface seems shallower than it really is; and to an observer looking across the pool from one side, it appears even shallower. The effect is caused by the refraction or bending of light.

Refraction occurs where light crosses a boundary and enters a different medium in which it travels with a different velocity

In this chapter we study refraction at plane surfaces, beginning with an experiment in which rays pass from air into glass.

16.2 REFRACTION AT AN AIR-GLASS INTERFACE

Fig. 16.1 illustrates the apparatus used for plotting the path of a ray through a thick semicircular block of glass. The straight edge of the glass is positioned on a line drawn on the paper, and a ray is directed towards a point halfway along the vertical plane face of the block. The paths of the rays are marked on the paper. Fig. 16.2 illustrates a typical set of rays. The

fig 16.1 apparatus used for plotting the path of a ray incident at a plane glass surface. The glass block has a semicircular section

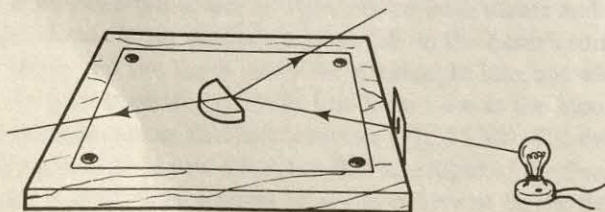
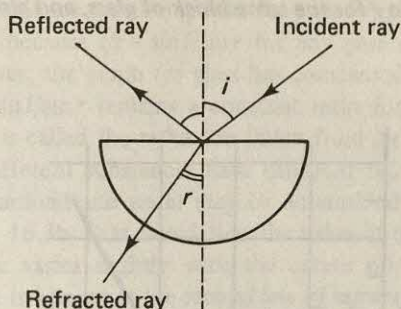


fig 16.2 *three rays and the normal are drawn where an incident ray meets the plane glass surface*



following properties may be observed:

- (i) the incident ray is partially reflected,
- (ii) refraction occurs where the ray enters the glass,
- (iii) the incident ray and the refracted ray are on opposite sides of the normal to the surface at the point of incidence,
- (iv) the plane of the figure includes all three rays as well as the normal at the point of incidence.

The observations (iii) and (iv) above illustrate the *first law of refraction*, which may be stated as follows:

The refracted ray lies in the plane formed by the incident ray and the normal to the surface at the point of incidence. The two rays are on opposite sides of the normal

The readings shown in Table 16.1 were obtained by measuring the angle of incidence i , the angle of refraction r , and then by repeating the experiment for different angles of incidence.

Table 16.1

Angle of incidence, i (deg)	72	61.5	50	32.5	23.5	11.5
Angle of refraction, r (deg)	38	35	30	20.5	15	7.5

Fig. 16.3a is a graph of the angle of incidence i plotted against the angle of refraction r . Three details should be noted. First, the lower portion of the graph is a straight line through the origin having a constant slope of i/r . Secondly, the slope increases at larger angles of incidence; and, thirdly, angles of incidence from 0° to 90° give rise to angles of refraction from 0° to only about 42° .

fig 16.3 (a) angles of incidence i , plotted against the corresponding angles of refraction r , for a ray entering glass from air. (b) $\sin i$ plotted against $\sin r$ for the same block of glass, and also for water and diamond

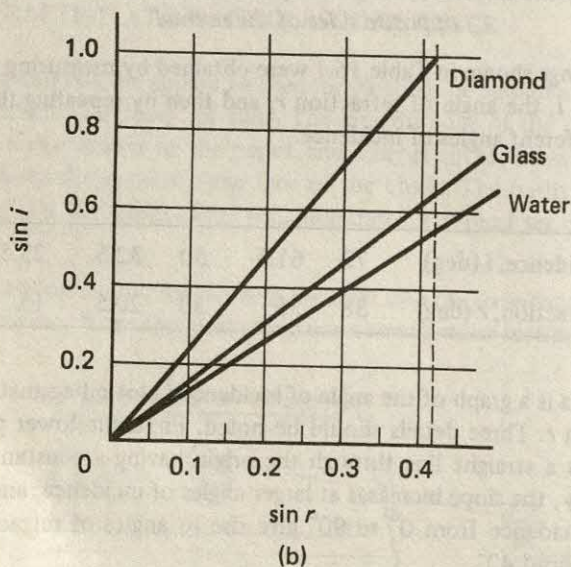
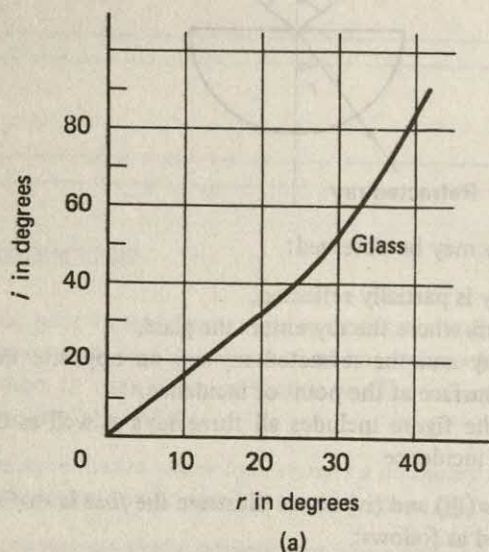


Fig. 16.3b shows the graph of $\sin i$ plotted against $\sin r$ for the same experiment, and also for similar experiments with water and diamond. At small angles of incidence the two graphs for glass (Figs. 16.3a and b) have the same slopes, because $i/r = \sin i/\sin r$ for any pair of small angles. In Fig. 16.3b, however, the graph for glass has constant slope throughout its entire range; so $\sin i/\sin r$ remains a constant ratio for all angles of incidence. The ratio is called the *refractive index* from air to glass and given the symbol n . Different substances have different refractive indices; the values of n for diamond and water may be determined from the slopes of the graphs in Fig. 16.3b. It is found that the value of the refractive index n for a substance varies slightly with the colour of the incident light. *Snell's law*, which is known as the *second law of refraction*, may be stated as follows:

When a ray of light of one colour crosses the boundary from medium 1 to medium 2, the ratio $\sin i/\sin r$ is a constant called the refractive index from medium 1 to medium 2, for light of that colour

In symbols,

$$n = \frac{\sin i}{\sin r}$$

16.3 REFRACTIVE INDICES

The *absolute refractive index* of a substance is the ratio $\sin i/\sin r$ when the incident ray is in a vacuum. Its value is very nearly the same as the refractive index from air to the substance. When a refractive index is given for a different boundary, the media on each side of the boundary must be named. Thus ${}_{\text{water}}n_{\text{glass}} = 1.12$ states that for a ray passing from water into glass, the refractive index is 1.12. Some refractive indices of substances in air are given in Table 16.2.

The composition of glass can be varied according to the use to be made of it, and its refractive index varies with the composition. Diamond has a very high refractive index, and it is said to be 'optically more dense' than

Table 16.2

<i>Substance</i>	<i>Refractive index in air</i>
Glass	1.50-1.56
Water	1.33
Carbon bisulphide	1.63
Diamond	2.42

substances whose refractive indices are lower. Optical density differs from the mass per unit volume of a substance; and a substance of lower density does not always have a smaller refractive index.

In Section 14.9(b) the ratio $\sin i/\sin r$ for waves was shown to equal the ratio

$$\frac{\text{Wave velocity in medium 1}}{\text{Wave velocity in medium 2}}$$

So the refractive index ${}_1n_2$ is equal to v_1/v_2 , where v_1 and v_2 are the wave velocities in medium 1 and medium 2 respectively:

$${}_1n_2 = \frac{\sin i}{\sin r} = \frac{v_1}{v_2} \quad (16.1)$$

The velocity of light in a substance can be calculated from the refractive index of the substance in air, since the speed of light in air is known. For example, the velocity of light in glass of refractive index 1.5 can be found using equation (16.1), given that the velocity of light in air is 3×10^8 m/s:

$$\begin{aligned} {}_{\text{air}}n_{\text{glass}} &= v_{\text{air}}/v_{\text{glass}} \\ 1.5 &= 3 \times 10^8/v_{\text{glass}} \\ v_{\text{glass}} &= 3 \times 10^8/1.5 \\ &= 2 \times 10^8 \text{ m/s} \end{aligned}$$

The velocity of light is least in substances of highest refractive index.

16.4 THE REVERSIBILITY OF A RAY OF LIGHT

If the glass block in Fig. 16.1 having a semicircular cross-section is replaced by a parallel-sided glass block, then the emergent ray is parallel to the incident ray but laterally displaced from it. The paths of the incident and emergent rays can be marked with pins at A, B, C and D illustrated in Fig. 16.4, after which the lamp and slit can be removed. The pins are not really in line, but they appear so when viewed through the glass block in either direction. Rays can travel each way along this path.

The path of a ray of light is reversible

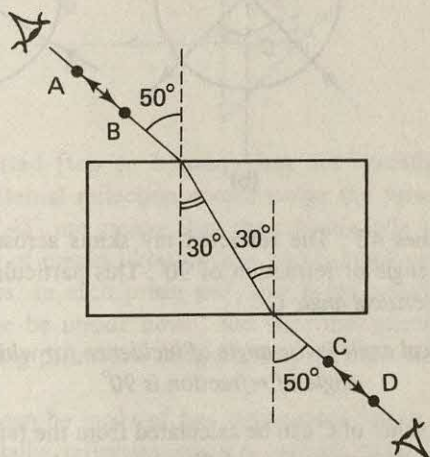
The refractive index from air to glass can be found from measurements of the angles of incidence and refraction at the first face of the glass. Similar measurements at the second face enable the refractive index from glass to air to be found. The results are:

$${}_{\text{air}}n_{\text{glass}} = \frac{\sin 50^\circ}{\sin 30^\circ} = 1.53$$

and

$${}_{\text{glass}}n_{\text{air}} = \frac{\sin 30^\circ}{\sin 50^\circ} = \frac{1}{1.53}$$

fig 16.4 a ray can pass in either direction along the same path



One refractive index is the reciprocal of the other. In general:

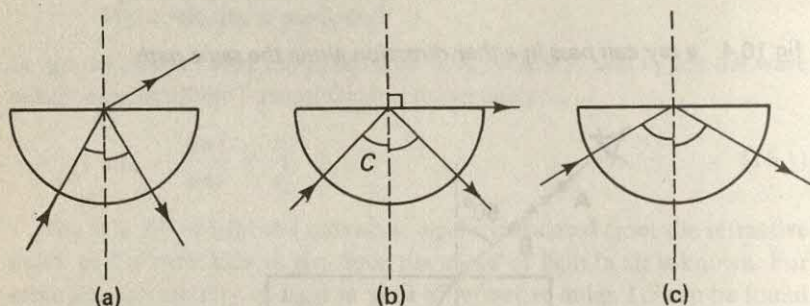
$${}_1n_2 = \frac{1}{{}_2n_1} \quad (16.2)$$

The result agrees with the expression for refractive index given in equation (16.1), where ${}_1n_2 = v_1/v_2$; but ${}_2n_1 = v_2/v_1$.

16.5 INTERNAL REFLECTIONS

A ray in air meeting a glass surface is partially reflected (Section 16.2). What happens when the incident ray in glass meets the glass-air surface from the other side? Fig. 16.5 illustrates three rays making different angles of incidence. In Fig. 16.5a where the angle of incidence is small, there is a refracted ray in air as well as a partially reflected ray which remains in the glass. Fig. 16.5b illustrates the same rays as the angle of

fig 16.5 (a) an incident ray making a small angle of incidence at a glass-air interface gives rise to reflected and refracted rays. (b) When the angle of incidence inside the glass reaches the critical angle, C , the angle of refraction is 90° . (c) Total internal reflection occurs when the angle of incidence inside the glass is greater than the critical angle



incidence approaches 42° . The refracted ray skims across the surface of the glass with an angle of refraction of 90° . This particular angle of incidence is called the *critical angle* C .

The critical angle is the angle of incidence for which the angle of refraction is 90°

The numerical value of C can be calculated from the refractive index of the glass. Imagine the rays in Fig. 16.5b are reversed, so that light passes from air into glass where $_{\text{air}}n_{\text{glass}} = 1.5$. Then $\sin 90^\circ / \sin C = 1.5$ and therefore $\sin C = \sin 90^\circ / 1.5$. $\sin C = 1/1.5 = 0.67$, from which the angle $C = 42^\circ$. In general, if n is the refractive index in air of a substance whose critical angle is C , then:

$$\sin C = 1/n \quad (16.3)$$

Fig. 16.5c illustrates the result of increasing the angle of incidence beyond the critical angle in glass. There is no refracted ray, and all the light is reflected. This condition is called *total internal reflection*.

Total internal reflection occurs when the angle of incidence is greater than the critical angle for a ray meeting the surface from the optically denser side

16.6 SOME APPLICATIONS OF TOTAL INTERNAL REFLECTION

The prism illustrated in Fig. 16.6 is a thick glass block having a cross-section in the form of an isosceles right-angled triangle. The rays in the

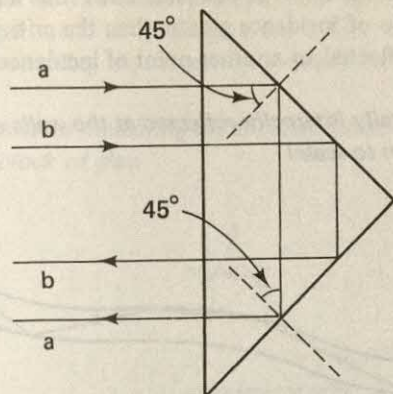
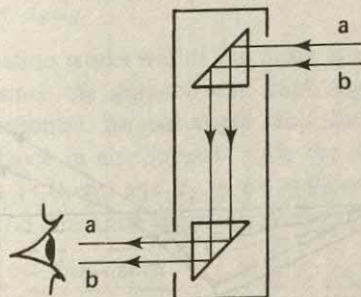
fig 16.6 *the reflecting prism inverts the rays a and b*

figure are inverted (top to bottom), but not laterally inverted (left to right). Total internal reflection occurs inside the prism since the angles of incidence of 45° are greater than the critical angle for a glass-air interface. Two pairs of prisms like this (one pair for each eye) are used in prismatic binoculars. In each prism pair, one prism inverts the image which would otherwise be upside down, and the other prism inverts the image laterally. By using prisms the designers have kept the instrument small and easy to handle.

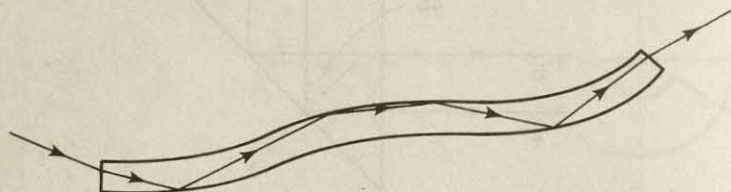
A periscope can be made of two right-angled prisms arranged as in Fig. 16.7. Their totally reflecting surfaces reflect more light than silvered mirror surfaces, and there is no problem of the silvered surfaces flaking off with age. For these reasons the periscopes of submarines incorporate reflecting prisms which need little or no maintenance, and have a much longer life than silvered mirrors.

Total internal reflection is important in fibre optics. Imagine a 'cable'

fig 16.7 *a periscope made with right-angled prisms needs no silvered surfaces*

of thin parallel glass fibres all closely bound together. What happens when light is admitted into the fibres at one end? Each time a ray meets the wall of a fibre at an angle of incidence greater than the critical angle, as in Fig. 16.8, it is totally reflected to another point of incidence further along the

fig 16.8 *a ray is totally internally reflected at the walls of a glass fibre.*
(Not drawn to scale)



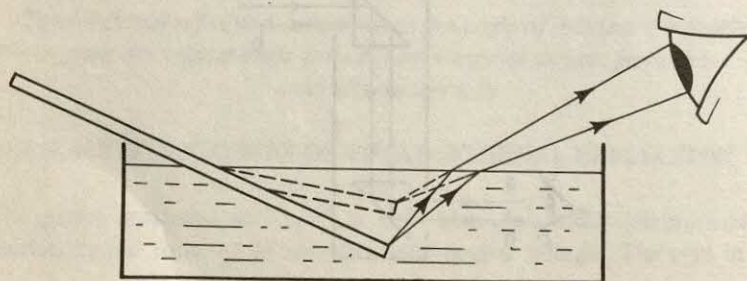
fibre. Here the same thing happens again, and so on until the ray emerges at the far end of the fibre. The fibres in the cable combine to provide a spy hole into the world at the far end. A picture is formed, made up of the information transmitted along the many fibres. Furthermore, the cable is flexible; it can carry its information around curves, but not where there is a sudden kink in it.

Optical fibres have been used in medicine, and more recently for communications purposes. The light passing along glass fibres can carry more information than electric currents in wires. It is possible to replace some of the many wires in a telephone network with fewer optical cables each of which can carry several conversations at the same time.

16.7 REAL AND APPARENT DEPTHS

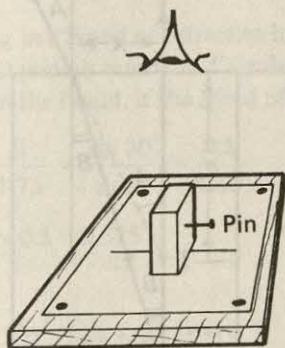
When a straight stick is dipped into water as illustrated in Fig. 16.9, the

fig 16.9 *a straight stick appears bent where it enters the water, because of refraction at the water's surface*



stick appears bent at the water's surface. The immersed parts of the stick appear closer to the observer's eye than they really are. Fig. 16.10 illustrates an experiment to investigate the effect, using a thick parallel-sided glass block standing on a ruled line. Viewed from directly above, the

fig 16.10 *apparatus for measuring the apparent thickness of a tall parallel-sided block of glass*



line appears closer to the top surface of the glass than it really is. Its virtual image is found by holding a horizontal pin above the line, and viewing both the pin and the line. When the pin is the same height above the paper as the line's image, the pin appears as an extension of the line; and movement of the observer's eye does not upset this alignment. Fig. 16.11 illustrates the position of the line at O, and its virtual image at I. Using the symbols in the figure, the real depth of the line d_r , and its apparent depth d_a , the refractive index from air to glass is:

$$\text{air } n_{\text{glass}} = \frac{\sin A}{\sin B} \approx \frac{x}{d_a} \div \frac{x}{d_r}$$

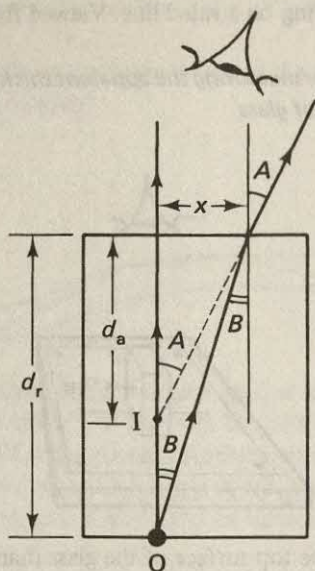
or

$$\text{air } n_{\text{glass}} = d_r/d_a \quad (16.4)$$

The approximation works well if the block is viewed along the normal to its top face (since the sines of very small angles A and B are almost equal to their tangents). We can apply the result to find the refractive index of a substance in air. Suppose d_r is the real thickness of a glass block and equals 11.0 cm, and d_a is the apparent thickness of the glass block and equals 7.3 cm. Then applying equation (16.4):

$$\text{air } n_{\text{glass}} = \frac{\text{Real depth}}{\text{Apparent depth}} = \frac{11.0}{7.3} = 1.51$$

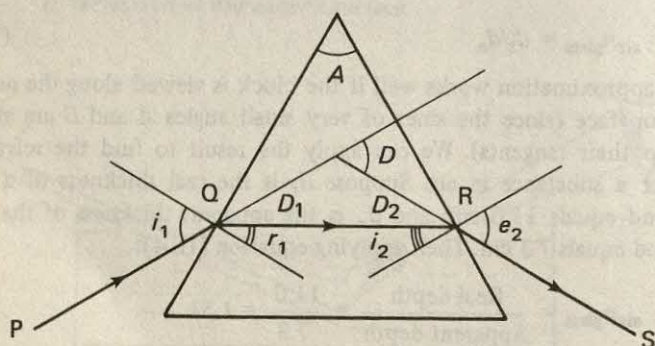
fig 16.11 rays from the line at O emerge from the block as if they had come from its virtual image at I



16.8 REFRACTING PRISMS

Fig. 16.12 illustrates the path of a ray through a 60° glass prism. The prism deviates the ray through an angle marked D in the figure. The *angle of deviation* D is the angle between the incident and emergent rays; and it is the sum of the separate deviations at the two surfaces of the prism. $D_1 = i_1 - r_1$, $D_2 = e_2 - i_2$, and the total deviation $D = D_1 + D_2$. The value

fig 16.12 a ray PQRS is deviated on entering and leaving a prism



of D for a given ray depends on the refractive index of the glass prism, which itself varies from one colour to another. It follows that the different colours which are combined in the ray are deviated by different amounts and separated from each other on passing through the prism. Refracting prisms are used to analyse the colours in beams of light emitted from various sources. This application is described more fully in Section 19.1.

WORKED EXAMPLE

A ray of light travelling in a liquid of refractive index 1.75 makes an angle of incidence of 30° at liquid/air interface. Calculate the angle of refraction and the speed of light in the liquid, if the speed in air is 3×10^8 m/s.

$$n_{\text{liquid}} n_{\text{air}} = \frac{1}{1.75} = \frac{\sin 30^\circ}{\sin r} = \frac{0.5}{\sin r}$$

$$\therefore \sin r = 1.75 \times 0.5 = 0.875$$

$$\therefore r = 61^\circ$$

$$n_{\text{air}} n_{\text{liquid}} = \frac{v_{\text{air}}}{v_{\text{liquid}}} \quad (\text{see page 152})$$

$$\therefore 1.75 = \frac{3 \times 10^8}{v_{\text{liquid}}}$$

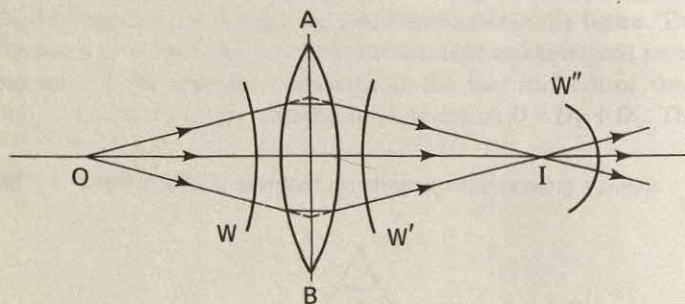
$$\therefore v_{\text{liquid}} = 1.7 \times 10^8 \text{ m/s}$$

THIN LENSES AND CURVED MIRRORS

17.1 ACTION OF A LENS

The passage of light through a thin converging lens is illustrated in Fig. 17.1. The light is refracted at both surfaces of the lens, but the rays emerge just as if they had suffered a single refraction at a plane somewhere inside the glass. For the purposes of ray constructions we represent a thin lens by a single line (AB in Fig. 17.1), where the refraction is considered to take place.

fig 17.1 *rays from O arrive at I as if they had been refracted at the plane AB inside the convex lens. The curves W, W' and W'' represent a sequence of wavefront positions*

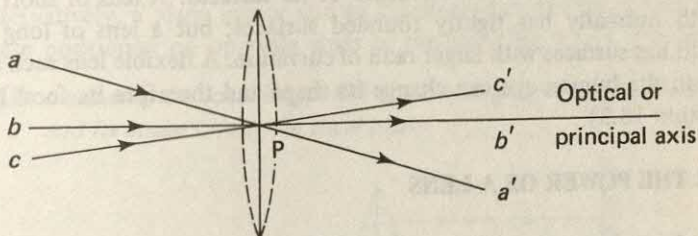


The curves labelled W, W' and W'' in Fig. 17.1 illustrate a sequence of wavefront positions for light waves spreading out from a point object at O, converging on its image position at I, and finally passing beyond the image position. The action of the lens is to change the shape of the wavefront, by slowing down the central portion so that it is overtaken by the outer edges. The central portion passes through the thickest part of the lens; and as light travels more slowly in glass than in air, the centre of the wave moves forward more slowly.

17.2 OPTICAL CENTRE AND FOCAL POINT

The axis of symmetry of a lens is called its *optical axis*, or *principal axis*. If we think of a thin lens as a single refracting plane, the point where the principal axis meets this plane is called the *optical centre*, or *pole P*, of the lens. Fig. 17.2 illustrates the passage of rays through this point. They pass

fig 17.2 rays *a, b* and *c* pass undeviated through the optical centre of the lens

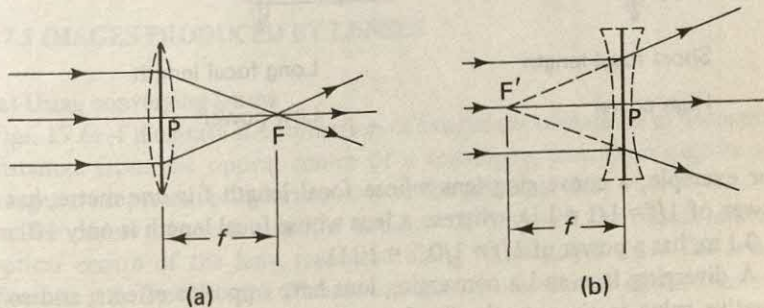


straight through the optical centre of the lens undeviated (this part of the lens behaves as if it were a thin pane of parallel-sided window glass).

The optical centre or pole of a thin lens is the point of the lens through which rays of light pass undeviated

Figs. 17.3a and b show how rays of light initially parallel to the principal axis pass through a convex lens and a concave lens respectively. In Fig. 17.3a the convex lens causes the rays to converge and pass through the point *F* on the principal axis. The convex lens is therefore a converging lens. *F* is called the *focal point* or *principal focus* of the lens. In Fig. 17.3b the concave lens causes the rays to diverge and appear to come from *F'*. The concave lens is a diverging lens. *F'* is its principal focus.

fig 17.3 (a) and (b) the focal length *f* of a lens is the distance between its optical centre and its focal point



The principal focus or focal point of a lens is the point towards which, or from which, light travels when incident rays parallel to its optical axis are refracted by the lens

The focal length f is the distance between the optical centre and the principal focus of a lens

The focal plane of a lens is the plane through the principal focus perpendicular to its optical axis

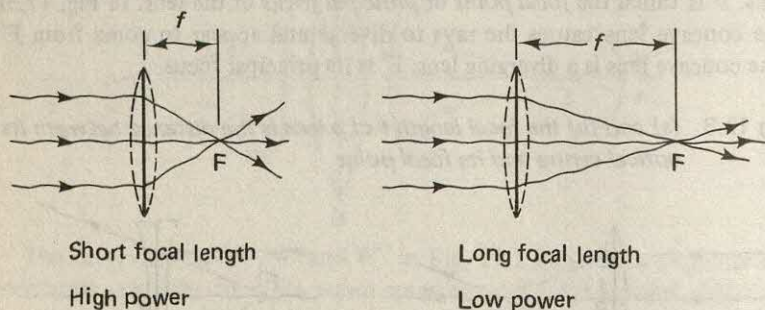
A rigid lens has a fixed focal length determined by the substance of the lens and the radii of curvature of its surfaces. A lens of short focal length normally has tightly rounded surfaces; but a lens of long focal length has surfaces with larger radii of curvature. A flexible lens such as the lens in the human eye can change its shape and therefore its focal length (Section 18.3).

17.3 THE POWER OF A LENS

The shorter the focal length of a lens, the more it converges or diverges the light passing through it. Power is defined so that a lens of short focal length has a higher power than one of long focal length (Fig. 17.4). The power is defined by the equation:

$$\text{Power in dioptres (D)} = \frac{1}{\text{Focal length in metres}} \quad (17.1)$$

fig 17.4 a lens of short focal length bends the light more, and has a higher power than a lens of long focal length



For example, a converging lens whose focal length f is one metre, has a power of $1/f = 1/1 = 1 \text{ D}$, whereas a lens whose focal length is only 10 cm, or 0.1 m, has a power of $1/f = 1/0.1 = 10 \text{ D}$.

A diverging lens and a converging lens have opposite effects, and so a negative value is given to the focal length and power of a diverging lens.

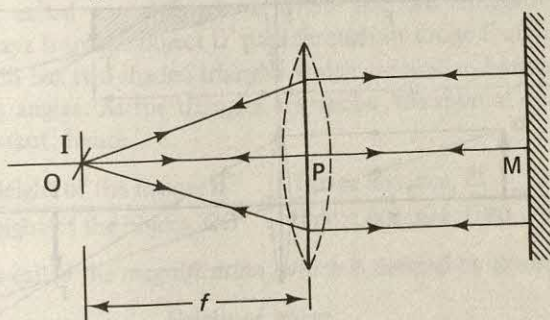
A lens whose principal focus lies 20 cm from its optical centre has a focal length of -0.2 m, and a power of $1/f = 1/(-0.2) = -5$ D.

17.4 MEASURING THE FOCAL LENGTH OF A CONVERGING LENS

Rays from a bright object a long way from a converging lens give rise to a real image in the focal plane of the lens. An approximate value for the focal length of the lens can be obtained by focusing the image of the distant object on a screen. The distance from the lens to the screen is the focal length of the lens (see Fig. 17.6a).

Alternatively, a plane mirror and an illuminated cross-wire mounted on card are positioned on opposite sides of the lens (Fig. 17.5). The object

fig 17.5 *measuring the focal length OP of a converging lens. The object O and its image I lie in the focal plane*



distance OP from the centre of the lens is adjusted so that the object and its image coincide. Rays from the centre of the cross form a parallel beam after passing through the lens. Each part of the beam meets the mirror normally, and is reflected back to its source along its original path. In this adjustment the distance OP is the focal length of the lens, since rays parallel to the optical axis are refracted towards the principal focus. The distance PM from the lens to the plane mirror is not important, and it can be any convenient distance.

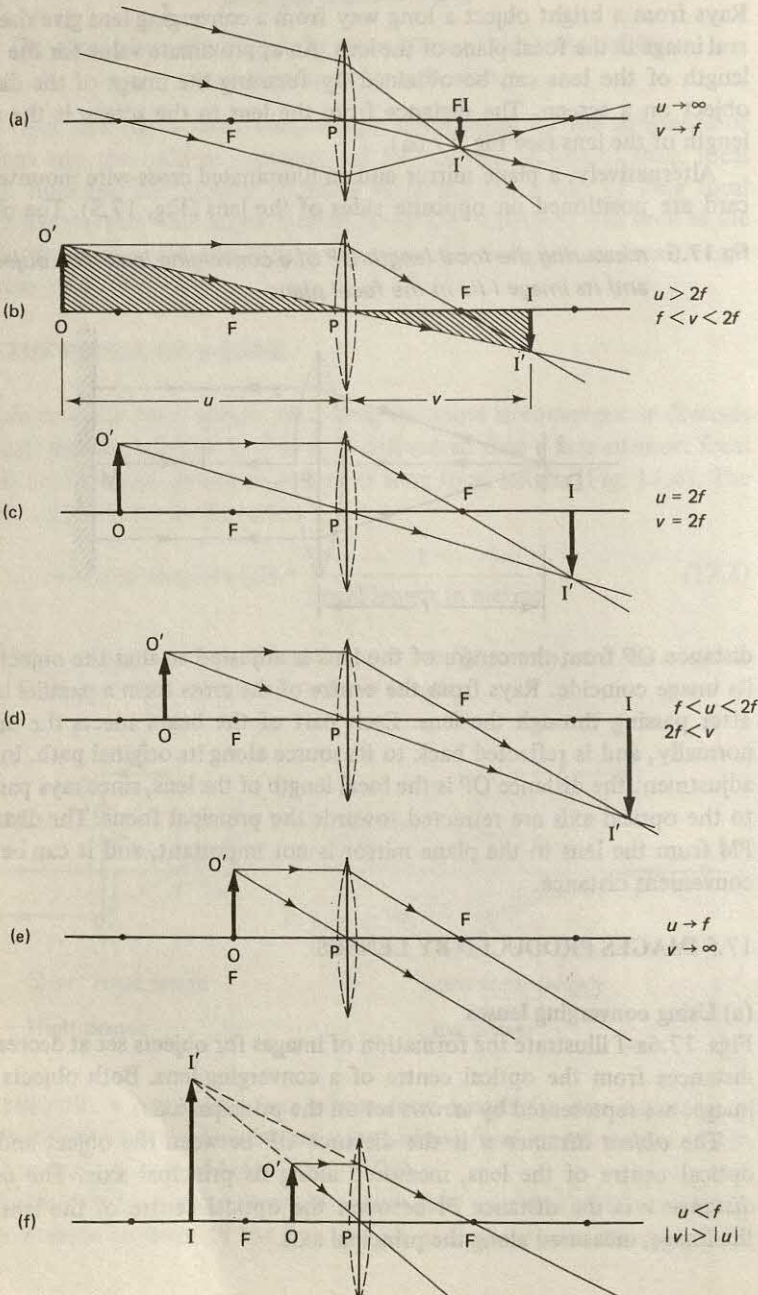
17.5 IMAGES PRODUCED BY LENSES

(a) Using converging lenses

Figs. 17.6a-f illustrate the formation of images for objects set at decreasing distances from the optical centre of a converging lens. Both objects and images are represented by arrows set on the principal axis.

The *object distance* u is the distance OP between the object and the optical centre of the lens, measured along its principal axis. The *image distance* v is the distance PI between the optical centre of the lens and the image, measured along the principal axis.

fig 17.6 as the object OO' gets closer to the converging lens, the image moves further from the lens and gets bigger. When the object is closer than one focal length from the lens, the image is virtual, magnified and erect



Both object and image distances can vary, but the focal length PF remains constant for a particular lens. In each figure there are dots on the principal axis at intervals of f (the focal length), measured from the optical centre on each side of the lens.

In Fig. 17.6a rays from a distant object point, which is above the axis, form a parallel incident beam which is brought into focus in the focal plane of the lens. The central ray passes through the optical centre without deviating (see Fig. 17.2).

In the remaining figures, the images are located by drawing:

- (i) the ray from O' which passes through the optical centre of the lens undeviated, and
- (ii) the ray which is initially parallel to the principal axis, and which passes through the principal focus after refraction.

The image, called a *real* image, is where the two emergent rays cross. Indeed all rays from the object O' pass through its image I' after refraction.

Fig. 17.6b has two shaded triangles which are similar, having three pairs of matching angles. As the triangles are similar, the ratio of corresponding sides is constant. Hence

$$\frac{\text{Height of the image, } II'}{\text{Height of the object, } OO'} = \frac{\text{Image distance, } PI}{\text{Object distance, } OP} = \frac{v}{u} \quad (17.2)$$

This ratio is called the *magnification*, which is defined by the equation:

$$\text{Magnification} = \frac{\text{Height of image}}{\text{Height of object}} \quad (17.3)$$

Equation (17.2) enables the magnification to be calculated if the object and image distances are known.

In Fig. 17.6c the object and image distances are equal when $u = v = 2f$, and the magnification is therefore one. If the object distance is made less than $2f$, the magnification exceeds one (Fig. 17.6d).

In Fig. 17.6e the object is one focal length from the lens, and the image is said to be *at infinity*. This is the reverse of the situation in Fig. 17.6a, in which the object is at infinity.

In Fig. 17.6f the object distance is made less than one focal length f , and a new situation arises. No real image is formed on the far side of the lens. Instead the rays emerge as if coming from I' . II' is a *virtual* upright image, whereas the previous images in Figs. 17.6a-e were all real and inverted. The importance of the arrangement in Fig. 17.6f is that the lens is being used as a magnifying glass to make things appear bigger than they are. Table 17.1 summarises the information gained from ray constructions for converging lenses and concave mirrors. As we shall see in Section 17.10, the same properties obtain for convex lenses as for concave mirrors.

Table 17.1 details of images for convex lenses and concave mirrors

Object distance, $OP = u$	Image distance, $PI = v$	Magnification, $m = \frac{PI'}{OO'} = \frac{v}{u}$	Type of image
$u \rightarrow \infty$	$v \rightarrow f$	∞	Real, inverted
u between $2f$ and ∞	v between f and $2f$	$m < 1$	Real, inverted
$u = 2f$	$v = 2f$	$m = 1$	Real, inverted
u between f and $2f$	v between $2f$ and ∞	$m > 1$	Real, inverted
$u \rightarrow f$	$v \rightarrow \infty$	∞	—
u less than f	Image distance greater than object distance	$m > 1$	Virtual, upright

*The term 'magnification' is not used when either the object or its image is at infinity.

(b) Using diverging lenses

Fig. 17.7 illustrates ray constructions for a diverging lens. The principal focus is a virtual point, and a ray initially parallel to the optical axis diverges as if coming from F' . The broken construction line is called a *virtual ray*. It establishes the virtual image position at I' . Whatever the object distance from a diverging lens, the image is always virtual, upright, diminished and closer to the lens than the object distance. Table 17.2 summarises the details of images produced by diverging lenses and convex mirrors. As we shall see in Section 17.10(b) the same properties obtain for concave lenses as for convex mirrors.

fig 17.7 a diverging lens produces a virtual, diminished and erect image

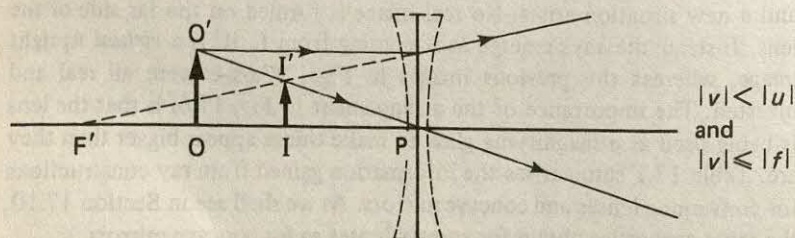


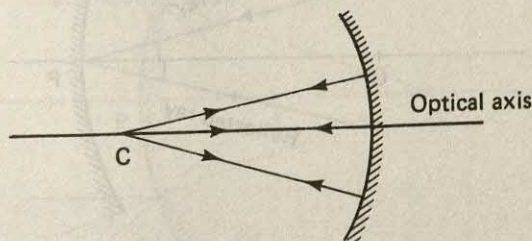
Table 17.2 *details of images for concave lenses and convex mirrors*

<i>Object distance,</i> $OP = u$	<i>Image distance,</i> $PI = v$	<i>Magnification,</i> $m = \frac{II'}{OO'} = \frac{v}{u}$	<i>Type of image</i>
Any distance up to infinity	v always a shorter distance than u or f	$m < 1$	Upright and virtual

17.6 THE ACTION OF A CURVED MIRROR

Rays reflected at curved mirrors obey the laws of regular reflection (Section 15.2). Small spherical mirrors are represented by the arcs of circles, each having a centre of curvature C , which lies on its axis of symmetry or optical axis (Fig. 17.8). Any straight line from C to the surface of a concave mirror is a normal to it, and so rays from C to the mirror return to C again. The centre of curvature and its image in the mirror thus coincide.

fig 17.8 *an object at the centre of curvature of a concave mirror coincides with its own image*



In Fig. 17.9 rays from an object point O are reflected so that the angle of incidence is equal to the angle of reflection. They pass through the image position I . The curves labelled W , W' and W'' illustrate a sequence of wavefront positions for light waves emitted from O (W), after subsequent reflection at the mirror (W'), and finally after passing through the image position (W''). Before reflection the wavefronts are centred on O , and after reflection on I . The mirror changes both the curvature and the centre of curvature of the wavefronts.

17.7 THE PRINCIPAL FOCUS AND FOCAL LENGTH

In Fig. 17.10 the optical axis meets a concave mirror at the optical centre or pole P , where it bisects the angle between incident and reflected rays.

fig 17.9 rays from O obey the laws of reflection and pass through I . The curves W , W' and W'' represent a sequence of wavefront positions for light waves before and after reflection

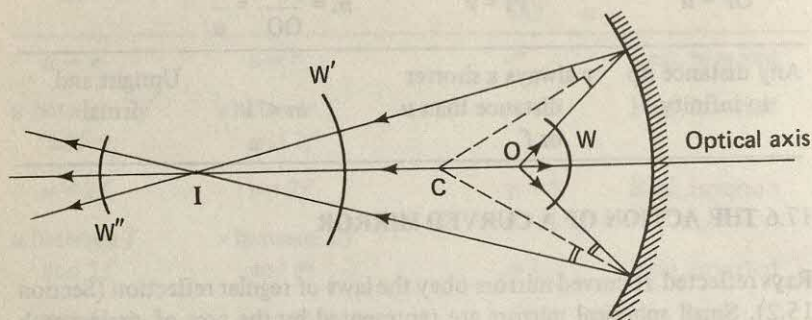
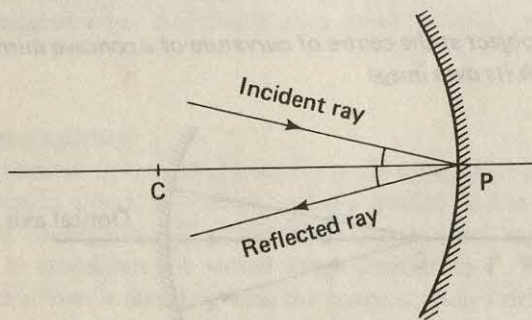


fig 17.10 an incident ray is reflected at the pole of a curved mirror, obeying the equal-angles law of reflection



Figs. 17.11a and b illustrate the reflection of rays which are initially parallel to the optical axes of a concave mirror and a convex mirror respectively. In Fig. 17.11a the rays converge and pass through the point F on the optical axis; and in Fig. 17.11b the rays diverge as if they had come from F' . F is called the *principal focus* (or *focal point*) of the concave mirror, and F' is the *principal focus* of the convex mirror. The distances FP and PF' are respectively the focal lengths of the mirrors.

The focal length of a curved mirror is the distance from the principal focus to the mirror, measured along its principal axis

The focal length of a curved mirror depends only on the curvature of its surface. In Fig. 17.11a triangle CFQ is isosceles, and therefore $CF = FQ$. For points of incidence Q close to the pole P , FQ is very nearly equal to the focal length FP . So for rays close to the principal axis:

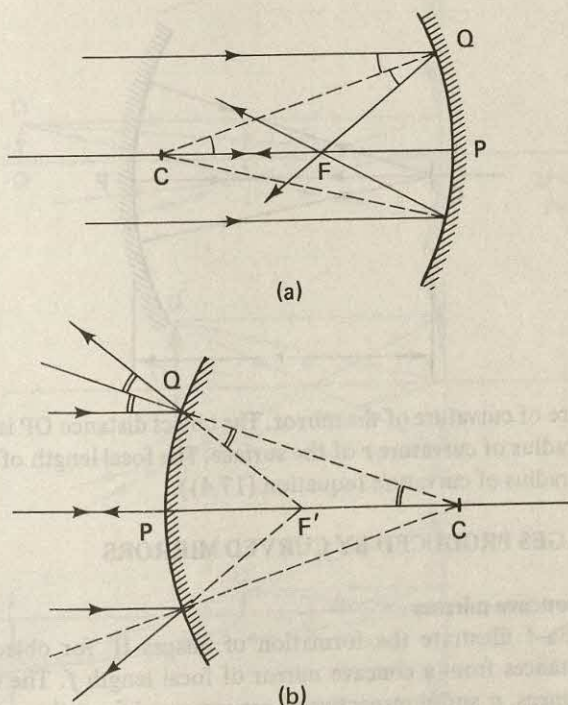
Radius of curvature $r = CP =$ Twice the focal length FP

or

$$r = 2f \quad (17.4)$$

A similar argument holds for the convex mirror (Fig. 17.11b).

fig 17.11 (a) and (b) the focal point of a curved mirror lies midway between the centre of curvature C and the pole P of the mirror



17.8 THE POWER OF A MIRROR

The power of a mirror is defined as for a lens (equation (17.1)):

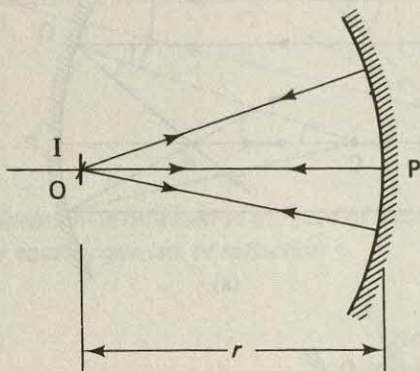
$$\text{Power in dioptres (D)} = \frac{1}{\text{Focal length in metres}}$$

As with lenses, mirrors of short focal lengths have greater powers than those of long focal length. For a concave mirror which causes a parallel-sided beam of light to converge, the power is positive; but a convex mirror causes a parallel-sided beam to diverge, and its focal length and power are given negative values.

17.9 MEASURING THE FOCAL LENGTH OF A CONCAVE MIRROR

To measure the focal length of a concave mirror, an illuminated cross-wire mounted on a white card is used as an object. It is moved along the optical axis until the cross-wire and its image coincide (Fig. 17.12). Both are then

fig 17.12 *when the object and image coincide, OP is the radius of curvature of the concave mirror.*



at the centre of curvature of the mirror. The object distance OP is measured; this is the radius of curvature r of the surface. The focal length of the mirror is half the radius of curvature (equation (17.4)).

17.10 IMAGES PRODUCED BY CURVED MIRRORS

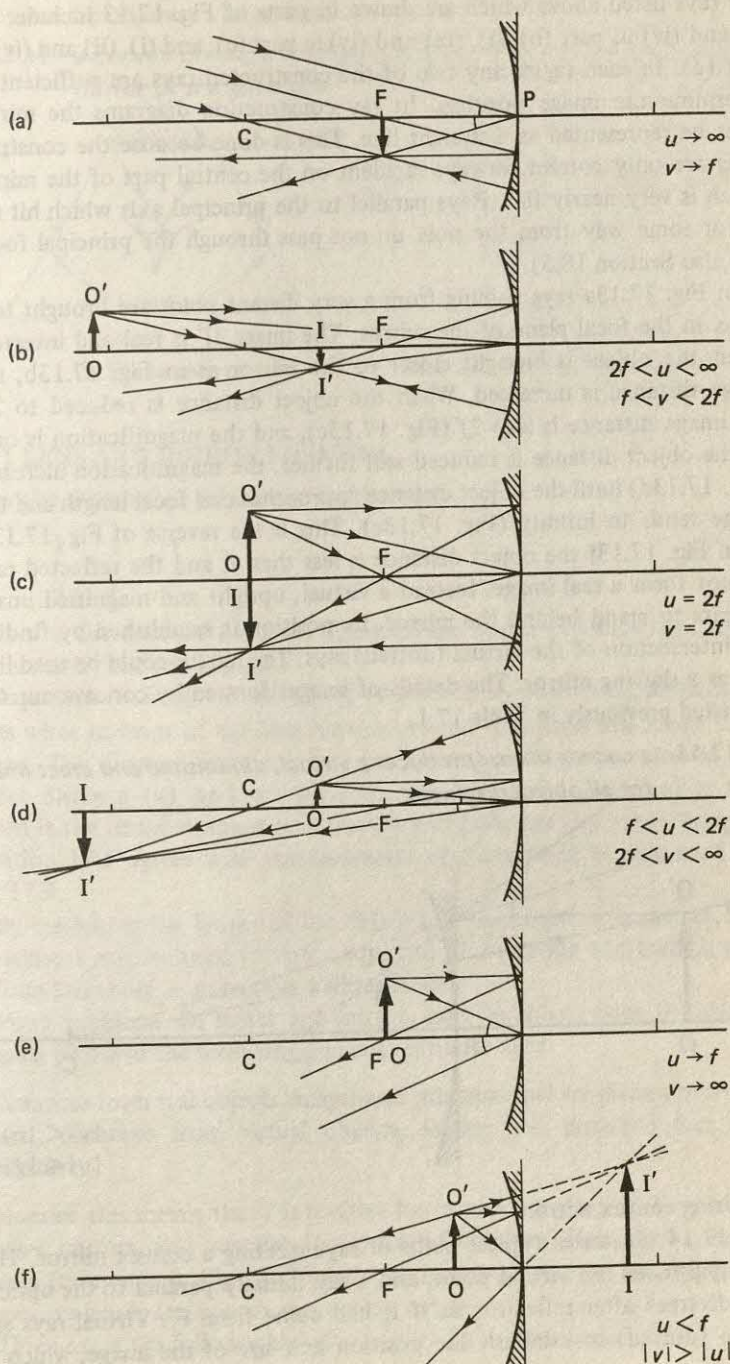
(a) Using concave mirrors

Figs. 17.13a-f illustrate the formation of images I' for objects OO' at various distances from a concave mirror of focal length f . The object and image distances, u and v respectively, are measured from the pole P of the mirror. The magnification of the image is defined in the same way in any optical system, and so the relationships (17.2) and (17.3) are valid for a mirror.

The image of O' (Figs. 17.13a-d) is the point through which all rays from O' pass after reflection at the mirror. The following construction rays may be drawn to determine the position of the image I' :

- (i) a ray parallel to the principal axis, which is reflected through the principal focus F ,
- (ii) a ray through the centre of curvature C , which meets the mirror normally and returns along its own path,
- (iii) a ray through the principal focus, which is reflected parallel to the principal axis,

fig 17.13 a concave mirror produces a real, inverted image whose magnification increases as the object distance is reduced. Finally when the object distance is less than one focal length, the image is virtual, magnified and erect



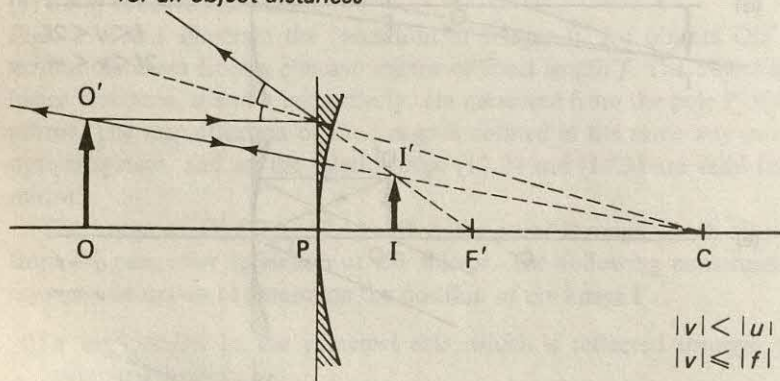
- (iv) a ray meeting the mirror at its pole, at which the principal axis is the normal—equal angles of incidence and reflection are constructed.

The rays listed above which are drawn in parts of Fig. 17.13 include: (i), (ii) and (iv) in part (b), (i), (iii) and (iv) in part (c), and (i), (ii) and (iv) in part (d). In each figure any two of the constructed rays are sufficient to determine the image position. In ray construction diagrams the mirror must be represented as a straight line. This is done because the constructions are only correct for rays incident on the central part of the mirror which is very nearly flat. Rays parallel to the principal axis which hit the mirror some way from the pole do not pass through the principal focus (see also Section 18.5).

In Fig. 17.13a rays coming from a very distant point are brought to a focus in the focal plane of the mirror. The image I' is real and inverted. When the object is brought closer to the mirror as in Fig. 17.13b, the image distance is increased. When the object distance is reduced to $2f$, the image distance is also $2f$ (Fig. 17.13c), and the magnification is one. As the object distance is reduced still further, the magnification increases (Fig. 17.13d) until the object distance approaches one focal length and the image tends to infinity (Fig. 17.13e). This is the reverse of Fig. 17.13a.

In Fig. 17.13f the object distance is less than f , and the reflected rays do not form a real image. Instead a virtual, upright and magnified image appears to stand behind the mirror. Its position is established by finding the intersection of the virtual (dotted) rays. The mirror could be used like this as a shaving mirror. The details of images formed by concave mirrors are listed previously in Table 17.1.

fig 17.14 *a convex mirror produces a virtual, diminished and erect image for all object distances*

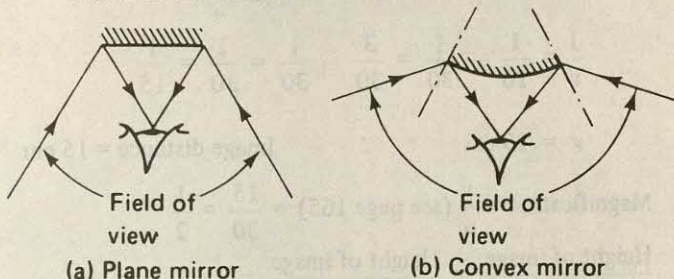


(b) Using convex mirrors

Fig. 17.14 illustrates typical paths of rays meeting a convex mirror. The principal focus is a virtual point, and a ray initially parallel to the optical axis diverges after reflection as if it had come from F' . Virtual rays are drawn (dotted) to establish the position and size of the image, which is

always closer to the mirror than is the object; the image is never more than one focal length from the mirror. The results are summarised previously in Table 17.2. This type of mirror is sometimes used for a driving mirror, as it gives the observer a larger field of view (see Fig. 17.15).

fig 17.15 a convex driving mirror gives a larger field of view than a plane mirror of the same size



17.11 LENS AND MIRROR FORMULA

It may be shown that for mirrors and lenses

$$\frac{1}{u} + \frac{1}{v} = \frac{1}{f} \quad (17.5)$$

where u is the distance from the object to the mirror or lens, v the image distance and f the focal length.

Equation 17.5 may be verified experimentally by positioning illuminated cross wires in front of the lens (or mirror) and forming a real image on a screen. The distance from the lens (or mirror) to the cross wires is the object distance (u), and the distance from the lens (or mirror) to the screen is the image distance (v). The value of f obtained by substituting in equation 17.5 agrees with measurements of f described in sections 17.4 and 17.9.

By measuring the height of the object and the height of the image, the experiment may be used to verify equation 17.2 (section 17.5) which will be found to apply to mirrors as well as lenses.

Many problems on lenses and mirrors may be solved using the above formula provided the following sign convention is used:

distances from real objects, images and principal foci are given a *positive* sign; distances from virtual objects, images and principal foci are *negative*.

In practice this means that f is positive for *convex* (converging) lenses and *concave* mirrors, and negative for *concave* (diverging) lenses and *convex* mirrors (reference to Figs 17.3b and 17.11b shows that the principal foci of concave lenses and convex mirrors are virtual).

Three examples follow which illustrate the use of equation 17.5:

Example 1 An object 2 cm high is placed 30 cm from a converging lens of focal length 10 cm. Where is the image and what is its height?

$$u = +30 \text{ cm} \quad f = +10 \text{ cm}$$

$$\frac{1}{u} + \frac{1}{v} = \frac{1}{f} \quad \therefore \quad \frac{1}{30} + \frac{1}{v} = \frac{1}{10}$$

$$\therefore \quad \frac{1}{v} = \frac{1}{10} - \frac{1}{30} = \frac{3}{30} - \frac{1}{30} = \frac{2}{30} = \frac{1}{15}$$

$$\therefore \quad v = 15 \text{ cm} \quad \therefore \quad \text{Image distance} = 15 \text{ cm}$$

$$\text{Magnification} = \frac{v}{u} \text{ (see page 165)} = \frac{15}{30} = \frac{1}{2}$$

$$\frac{\text{Height of image}}{\text{Height of object}} = \frac{\text{Height of image}}{2} = \frac{1}{2}$$

$$\therefore \quad \text{Height of image} = 1 \text{ cm}$$

Example 2 An object is placed 20 cm from a diverging lens of focal length 5 cm. Where is the image?

$$u = +20 \text{ cm} \quad f = -5 \text{ cm (note the negative sign for a diverging lens)}$$

$$\therefore \quad \frac{1}{+20} + \frac{1}{v} = \frac{1}{-5}$$

$$\therefore \quad \frac{1}{v} = \frac{1}{-5} - \frac{1}{20} = -\frac{4}{20} - \frac{1}{20} = -\frac{5}{20} = -\frac{1}{4}$$

$$\therefore \quad v = -4 \text{ cm}$$

The image is 4 cm from the lens and virtual.

Example 3 A concave mirror has a radius of curvature of 30 cm. A virtual image is formed 30 cm from the mirror. Where is the object?

Since the focal length is half the radius of curvature (page 169)

$$f = +15 \text{ cm} \quad v = -30 \text{ cm (negative since the image is virtual)}$$

$$\frac{1}{u} + \frac{1}{v} = \frac{1}{f} \quad \therefore \quad \frac{1}{u} + \frac{1}{-30} = \frac{1}{+15}$$

$$\therefore \quad \frac{1}{u} = \frac{1}{15} + \frac{1}{30} = \frac{3}{30} = \frac{1}{10}$$

$$u = 10 \text{ cm}$$

OPTICAL INSTRUMENTS

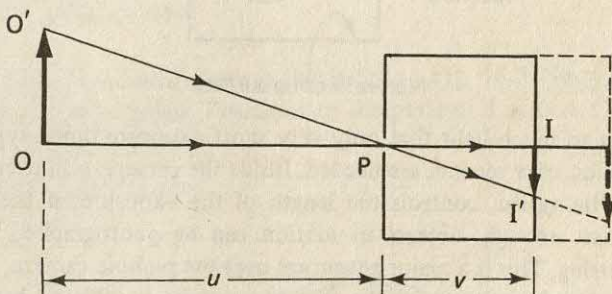
AND THE EYE

18.1 THE PINHOLE CAMERA

The pinhole camera is a simple device for taking photographs. Light is admitted into a box through a small pinhole in one side, and the image is formed on a film on the opposite side of the box (Fig. 18.1). As in other cameras, the box has dull black internal surfaces to prevent reflected light forming secondary images on the film.

Fig 18.1 shows the formation of a real, inverted image II' of an object OO' . The image would be taller if the box were longer, or if the camera

fig 18.1 *a real, inverted image is formed in a pinhole camera. The dotted lines show that the image is enlarged when the box is longer*



were held closer to the object. The magnification (equation (17.3)) is II'/OO' , which equals the ratio PI/OP or v/u in the similar triangles $II'P$ and $OO'P$. The object and image distances are measured from the pinhole P .

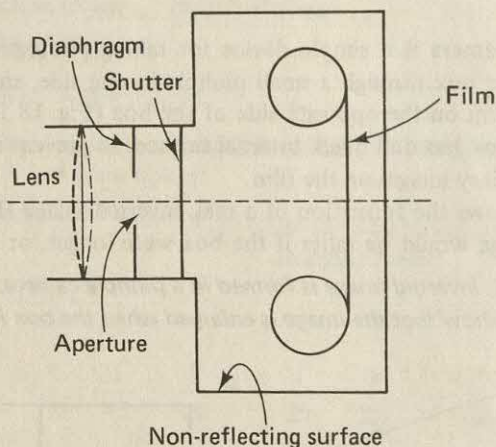
Provided the pinhole is small, the images of stationary objects at all distances appear reasonably in focus. This is an advantage not found in cameras which have lenses. The pinhole camera is also very cheap. Its main disadvantage is that the pinhole lets in very little light, and so a long

exposure time is needed for each picture. Only stationary objects may therefore be photographed. A way of reducing the exposure time is to use very sensitive film. The alternative of enlarging the pinhole is not satisfactory; for although it makes the image brighter, a larger pinhole causes the image to be unacceptably blurred.

18.2 THE CAMERA WITH A LENS

The camera illustrated in Fig. 18.2 has a converging lens to focus a real, inverted image on the film. Light enters the camera through its *aperture*, which is a hole of variable diameter in an adjustable *diaphragm*. The aper-

fig 18.2 a camera



ture lets in so much light that only very short exposure times, typically a small fraction of a second, are needed. Inside the camera, a shutter, whose speed can be varied, controls the length of the exposure; if the shutter speed is fast enough, objects in motion can be photographed without undue blurring. This is a major advantage over the pinhole camera.

Focusing a sharp image on the film in a camera is achieved by moving the lens in relation to the film. For very distant objects, the lens must be one focal length in front of the film; but for closer objects the image distance is greater, and the lens must be moved slightly further from the film.

In many photographs some foreground and background detail can be seen, giving the appearance of *depth* to the field.

The depth of field is the range of distances for which images are reasonably in focus

The depth of field can be increased by reducing the size of the aperture, so that light is admitted to the camera only through the central zone of its lens. The exposure time must be increased to compensate for the reduced illumination of the image, or the film is under-exposed.

On dull days the aperture is normally made larger, or the exposure time is increased, so that sufficient light reaches the film; but when photographs are taken on bright sunny days, the aperture can be made smaller without the need for long exposures. Some typical camera adjustments are summarised in Table 18.1.

Table 18.1 *some camera adjustments*

<i>Situation</i>	<i>Camera settings</i>
Distant object	Lens one focal length from the film
Close object	Lens slightly further from the film
Moving object	Very short exposure and so larger aperture
Dull day	Larger aperture or longer exposure or both
Bright day	Smaller aperture or shorter exposure or both
Greater depth of the field	Smaller aperture and so longer exposure

18.3 THE HUMAN EYE

A right eye seen from above is illustrated in Fig. 18.3. Some of its parts perform broadly similar functions to the parts of a camera (Table 18.2).

fig 18.3 *the human eye*

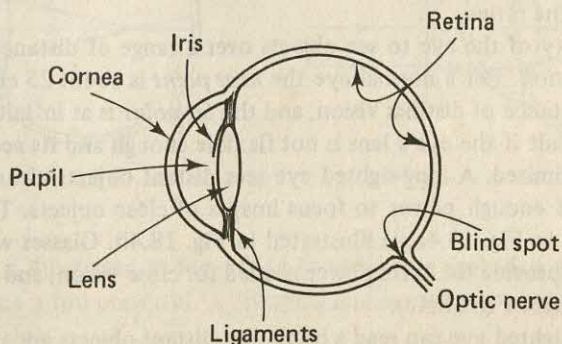


Table 18.2 *a comparison between the eye and the camera*

<i>Eye</i>	<i>Camera</i>
Real, inverted, diminished image	Real, inverted, diminished image
Non-reflecting surfaces	Non-reflecting surfaces
Retina	Light-sensitive film
Flexible lens	Rigid lens
Focused by reshaping the lens	Focused by moving the lens
Pupil	Aperture
Iris	Diaphragm

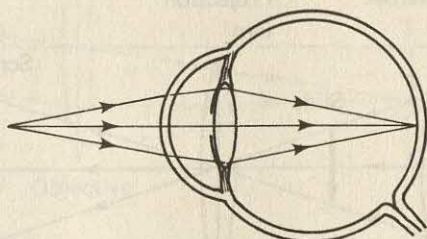
Both have a converging *lens* which forms a real, inverted image on a light-sensitive surface. The inside surfaces of both the eye and the camera are non-reflecting, and they both have a means of controlling the amount of light forming the image. In the eye, sensitive nerve endings respond to the image formed on the *retina* and transmit impulses along the *optic nerve* to the brain. At the place where the optic nerve leaves the eye, there is an insensitive spot on the retina called the *blind spot*.

The eye's *pupil* is the hole which lets in the light just as the aperture does in a camera. The pupil's size is controlled by the coloured *iris* which surrounds it, causing the pupil to dilate or enlarge in a darkened room, and to shrink when bright lights are turned on. The *cornea* is the curved transparent material at the front of the eye where rays experience some refraction. They are refracted further by the eye's lens, which focuses the image on the retina. The lens is flexible and is shaped by muscles around its edge. The power and focal length of the lens vary as the curvature of its surfaces alters. In this way the images of objects at various distances can be sharply focused on the retina.

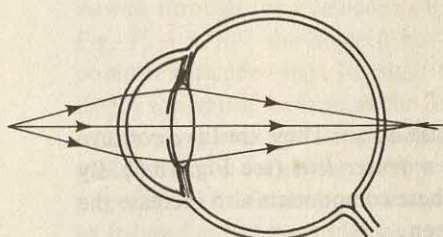
The ability of the eye to see objects over a range of distances is called *accommodation*. For a normal eye the *near point* is about 25 cm away, at the least distance of distinct vision, and the *far point* is at infinity. Defects of vision result if the eye's lens is not flexible enough and its accommodation is too limited. A long-sighted eye sees distant objects clearly, but its lens has not enough power to focus images of close objects. The defect, not present in Fig. 18.4a, is illustrated in Fig. 18.4b. Glasses with a converging lens provide the extra power needed for close vision, and the defect is corrected (see Fig. 18.4c).

A short-sighted eye can read a book, but distant objects appear blurred.

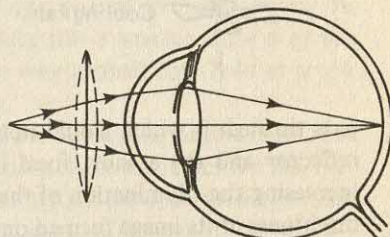
fig 18.4 (a) an eye with the image in focus on the retina. (b) Long sight. The eyeball is too short, or the lens not powerful enough to focus images of nearby objects. (c) Correction for long sight using a converging lens. (d) Short sight. Distant objects cannot be seen clearly. (e) Correction for short sight, using a diverging lens



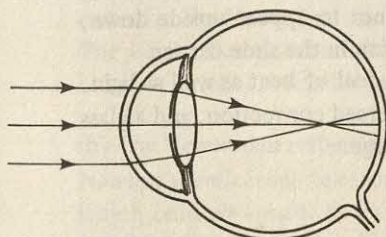
(a)



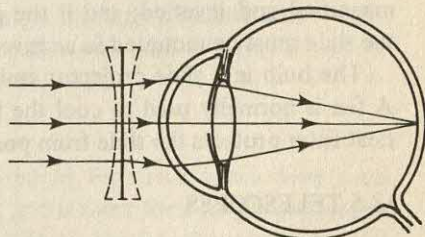
(b)



(c)



(d)



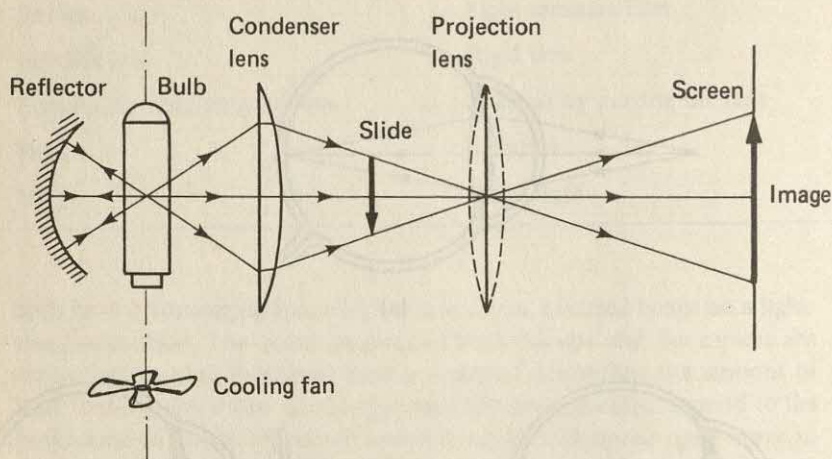
(e)

The defect is illustrated in Fig. 18.4d in which the eyeball is too long, or the eye's lens is too powerful. A diverging lens is used to correct the defect (see Fig. 18.4e).

18.4 THE SLIDE PROJECTOR

Fig. 18.5 illustrates the main features of the slide projector. Light from a bright filament lamp illuminates a small slide or transparency. Two features are included which increase the illumination of the slide by causing rays to

fig 18.5 the slide projector (not to scale)



pass through it which would not otherwise do so. They are (a) a *concave reflector* and (b) a short focal length *condenser lens* (see Fig. 18.5). By increasing the illumination of the slide, these components also increase the brightness of its image formed on the screen.

The projection lens has a focal length typically about 8 cm. It is positioned just more than one focal length from the slide, and adjusted so that the image is sharply focused on the distant screen. The image is real, magnified and inverted; and if the picture is not to appear upside down, the slide must be mounted in an inverted position in the slide carrier.

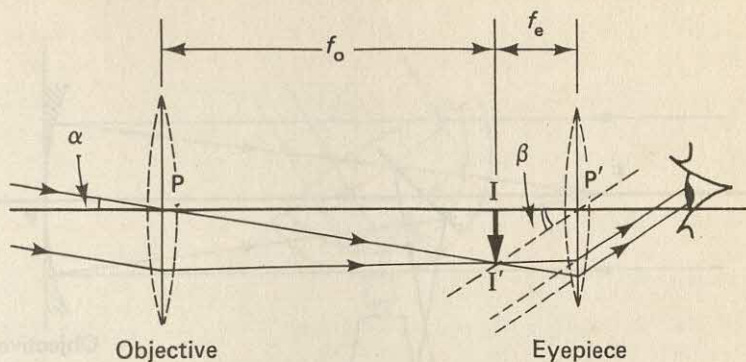
The bulb in a slide projector emits a good deal of heat as well as light. A fan is normally used to cool the bulb by forced convection, and a glass heat filter protects the slide from possible damage.

18.5 TELESCOPES

A telescope enables detail on distant objects to be seen more clearly.

(a) The astronomical refracting telescope

The telescope consists of two converging lenses. The *objective lens* has a long focal length, and the *eyepiece* has a short focal length. In Fig. 18.6 parallel rays from a point on a distant object are brought to a focus at I' .

fig 18.6 *the astronomical telescope*

A real, inverted image II' is formed in the focal plane of the objective lens. This image is much smaller than the object, but it appears larger when viewed through the eyepiece as it fills a larger part of the field of view. In Fig. 18.6 it fills the angle β but the object fills a smaller angle α at the observer's unaided eye. In much the same way a small coin held at arm's length can appear as large as the full moon.

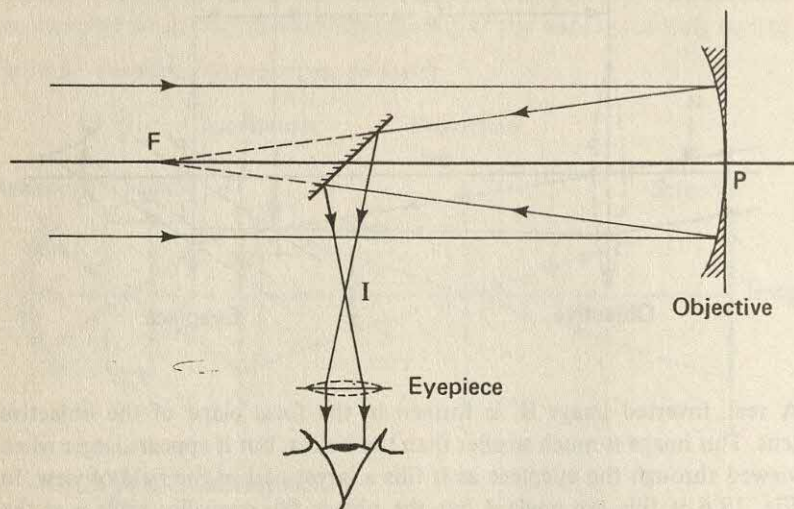
The eyepiece acts as a magnifying glass. In normal adjustment it is positioned one focal length from II' , and II' acts as an object for the eyepiece lens. The final virtual image is at infinity, so the observer's eye must be focused as if it were looking into the distance.

The astronomical telescope is unsuitable for viewing objects on Earth, as its final image is inverted; but this does not matter when stars or planets are viewed. The amount of detail revealed on a distant object, such as a crater on the Moon, increases when a larger diameter objective lens is used. The lens collects more light than a smaller one, and it forms a brighter image II' in which detail is more clearly defined.

(b) The Newtonian reflecting telescope

Newton's reflecting telescope illustrated in Fig. 18.7 has a long focal length concave mirror as its objective, and a short focal length converging eyepiece. The real image which would be formed at F in Fig. 18.7 is formed instead at I because a small plane mirror is positioned at 45° to the optical axis of the objective. The image at I can be examined without the observer having to sit in the path of the incident beam near F . Thus the principal axes of the objective and eyepiece are perpendicular. In normal adjustment the final virtual image is at infinity, and the eye's lens is relaxed for distant viewing.

fig 18.7 *Newtonian reflecting telescope. The real image would be at F, but the plane mirror reflects the rays so that the image is at I*



The brightness of the image at I and the amount of detail it reveals depend on the diameter of the objective mirror. The Newtonian telescope has a big advantage over the refracting telescope, as it is possible to support a very large diameter concave mirror by means of a framework behind the mirror's surface. Unfortunately a converging lens of large diameter is very heavy, and its shape becomes distorted as it can only be supported around its rim.

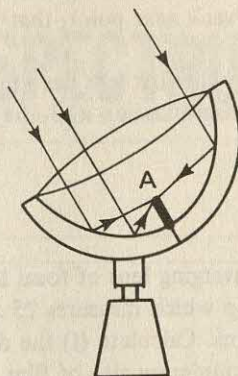
A large diameter concave objective mirror presents another problem. A large spherical mirror does not bring a parallel beam of light to a point focus. A parabolic mirror is needed to bring a wide beam of incident light to a point focus.

(c) Radio telescopes

Fig. 18.8 illustrates a typical radio telescope having a wide parabolic reflector. It is mounted so that its axis can be directed at any point in the sky. At the focus of the parabolic reflector there is a radio aerial, which only receives strong signals if the distant source lies on the principal axis of the 'dish' reflector. Radio signals received by the aerial are analysed electronically.

One of the values of a radio telescope, like the one at Jodrell Bank, is for detecting radio wave sources in the sky which may not be intense emitters of light. Radio stars are not all emitters of light.

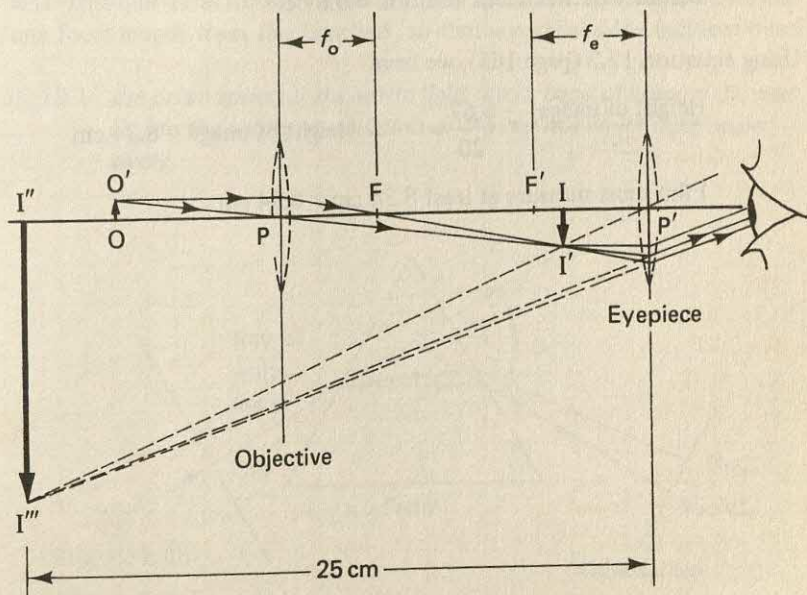
fig 18.8 a radio telescope has a concave parabolic reflector which brings radio waves to a focus at the aerial, A



18.6 THE COMPOUND MICROSCOPE

The compound microscope in Fig. 18.9 has a converging objective lens of very short focal length, and a short focal length eyepiece. The object distance OP is slightly greater than the objective's focal length, and a

fig 18.9 the compound microscope. The final magnified, virtual image is at the near point in normal adjustment



magnified, real image is formed at II' between the lenses. The eyepiece is positioned slightly less than one focal length from II' so as to produce a magnified, virtual image. The greatest magnification is produced when the final image is at the observer's near point, that is, about 25 cm from the eye.

If the image produced by the first lens has a magnification $\times 20$, and the eyepiece causes a further magnification $\times 10$, the total magnification $\times 200$ would be achieved.

WORKED EXAMPLE

A simple camera has a converging lens of focal length 5 cm. It is used to take a photograph of a map which measures 25 cm $\times$ 25 cm and is placed 20 cm from the camera lens. Calculate (i) the distance between the lens and the film and (ii) the minimum size of film that must be used if the whole of the map is to appear on the film.

Using the equation on page 172 we have

$$\frac{1}{u} + \frac{1}{v} = \frac{1}{f}$$

$$\therefore \frac{1}{+20} + \frac{1}{v} = \frac{1}{+5} \quad \therefore \frac{1}{v} = \frac{1}{5} - \frac{1}{20} = \frac{3}{20} \quad \therefore v = 6.67 \text{ cm}$$

Distance between lens and film = 6.67 cm

Using equation 17.2 (page 165), we have

$$\frac{\text{Height of image}}{25} = \frac{6.67}{20} \quad \therefore \text{Height of image} = 8.34 \text{ cm}$$

Film must measure at least 8.34 cm $\times$ 8.34 cm

THE ELECTROMAGNETIC SPECTRUM

19.1 VISIBLE LIGHT

Light of many colours can be derived from white light, e.g. from sunlight. A band of colours or a rainbow appears when direct sunlight is refracted by raindrops in the sky; and a similar band of colours is produced when a ray of white light from a tungsten filament lamp is refracted on passing through a triangular glass prism (Fig. 19.1). The band of colours displayed on the white screen is called a *spectrum*. It is not a *pure spectrum*, however, as some of the colours in the middle of the band overlap one another.

In Fig. 19.2 a converging lens before and after the prism enable each colour to be brought to a separate focus in the focal plane of the second lens. This illustrates the formation of a pure spectrum. The slit is positioned one focal length from the first lens, so that a parallel-sided incident beam

fig 19.1 *the prism splits up the white light into a band of colours. D_r and D_v are the angles of deviation of the red and violet light respectively*

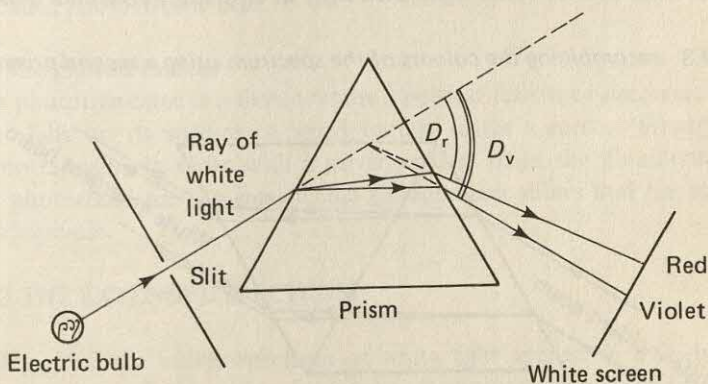
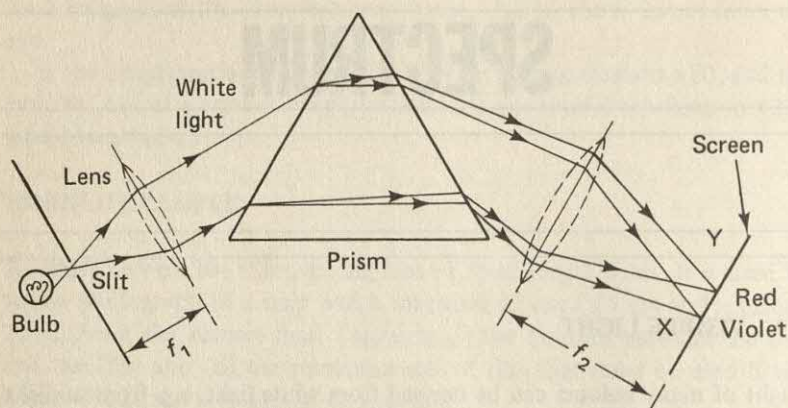


fig 19.2 *the production of a pure spectrum. The small angle between the emergent red and violet beams is called the dispersion*

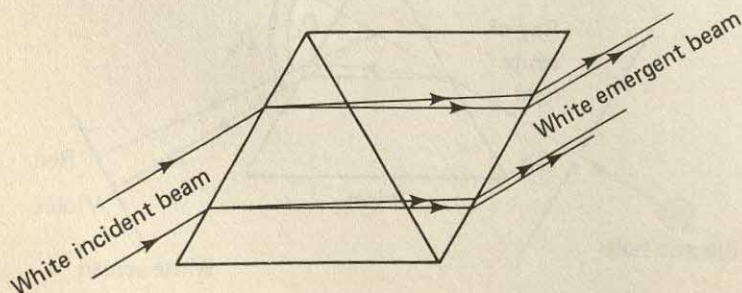


of white light meets the prism. The refractive index of glass (Section 16.3) depends on the colour of the light. It is slightly greater for violet light than for red, and so violet light is deviated by the prism more than red light. The screen is positioned one focal length beyond the second lens.

Sir Isaac Newton used a prism to split sunlight into a spectrum in which he recognised the colours red, orange, yellow, green, blue, indigo and violet arranged in that order. The colours formed a continuous band from red to violet. A second identical prism did not split the light into any new colours; but when the second prism was arranged as shown in Fig. 19.3, it recombined the colours to form an emergent beam of white light. This showed that white light is a combination of all the colours in the rainbow.

From the evidence of diffraction and interference experiments (Sections 14.9(c)(i) and 14.9(d)(ii)), the wavelength of red light is found to be greater than that of violet light. Red light in air has a wavelength close to

fig 19.3 *recombining the colours of the spectrum using a second prism*



750 nm, and violet light has a wavelength in air of about 420 nm (1 nm is 10^{-9} m). The complete spectrum of white light includes light of all wavelengths between the red and the violet.

19.2 DETECTORS OF LIGHT

(a) The eye

The retina at the back of the eye is a light-sensitive surface on which a real image is focused (Section 18.3). Nerve impulses from the retina are passed to the brain by way of the optic nerve. A normal eye can detect light of any wavelength in the range bounded by the wavelengths of red and violet light.

(b) The camera

In black-and-white photography, light is absorbed on reaching a photographic plate, and its energy triggers a chemical reaction. There is no reaction, however, where the plate receives no light energy. In colour photography, several chemical reactions are possible when light reaches the film. The ones which take place are determined by the colours of the incident light.

(c) The light meter

A photographic light meter is used to measure the light intensity before a photograph is taken. (It provides information which enables a photographer to adjust the aperture of the camera correctly.) Such meters can contain a surface which emits electrons when it receives light energy, and a microammeter which measures the flow rate of electric charge from the surface. The greater the light intensity, the greater is the current through the microammeter. The principle on which the meter works is called the *photoelectric* or *photoemissive effect*, and cells working on this principle are called *photoelectric cells*.

(d) The phototransistor

The phototransistor is a device whose electrical resistance decreases when light falls on its surface. A small battery drives a current through the phototransistor in series with a galvanometer. When the illumination of the phototransistor is increased, the galvanometer shows that the current also increases.

19.3 THE EXTENDED SPECTRUM

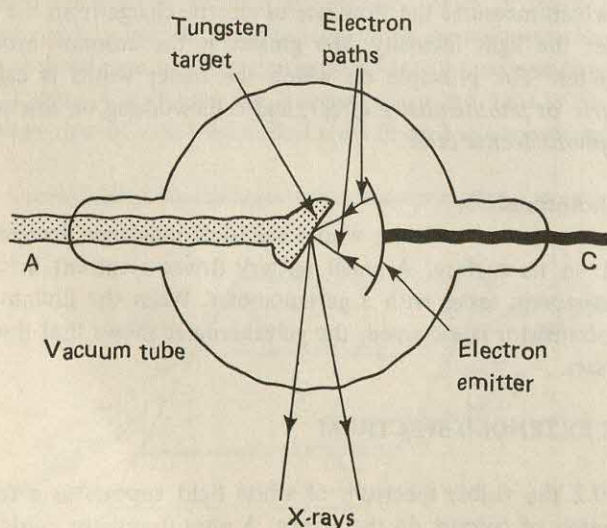
In Fig. 19.2 the visible spectrum of white light appears as a fairly well defined patch of colours on the screen. A phototransistor could be used

in this region to detect light rays of various colours; it also detects the presence of rays in the regions beyond the red and violet ends of the spectrum. Some photoelectric cells and camera films also show that there is light in these regions, although the eye cannot see it. Rays beyond the violet end of the spectrum at X are called *ultra-violet* rays, and those beyond the red end of the spectrum at Y are called *infra-red* (Fig. 19.2).

A mercury vapour lamp and a quartz lamp emit ultra-violet light as well as visible light. The Sun also emits ultra-violet light, though much of it reaching the Earth's atmosphere is absorbed before it reaches the ground. The Sun's ultra-violet rays cause a sunbather's skin to become suntanned; they also cause some substances to fluoresce. A fluorescent substance emits visible light when it absorbs ultra-violet light, so a fluorescent paper can be used to detect ultra-violet rays. Tests show that ultra-violet light of wavelength considerably less than that of violet light does not readily pass through glass. A window pane absorbs the ultra-violet while transmitting the incident visible light.

X-rays have shorter wavelengths in air than ultra-violet rays, and they are more penetrating. They are emitted when a tungsten target contained in a vacuum tube is bombarded by very fast-moving electrons (Fig. 19.4). They can be detected photographically or by their effect on a fluorescent screen. The rays penetrate the human body, but they are partially absorbed by

fig 19.4 an X-ray tube. A high voltage applied between the anode A and cathode C accelerates electrons towards the tungsten target. Their energies are converted into X-rays



bones which show up well as shadows in photographs. X-rays are therefore invaluable in the investigation of suspected bone fractures.

γ -rays (see also Chapter 28) have shorter wavelengths than X-rays, and they are even more penetrating than X-rays. They affect a photographic plate, and they can also be detected by the ionising effect that they have on molecules in their path. γ -rays are emitted from the nuclei of certain radioactive atoms. They possess great energy, and like X-rays they are harmful to the body. Their destructive effect on cell life can be put to good use, as γ -rays can destroy bacteria without harming a foodstuff in the same package.

Beyond the red end of the spectrum, *infra-red* light has wavelengths greater than red light. The rays are emitted by all hot objects, and they are referred to as *radiant heat*. The skin can feel the effect of radiant heat, which can also be detected using a phototransistor, or a thermometer with a blackened bulb. Glass tends to absorb radiant heat and block its flow, so that infra-red radiation does not readily pass out through the glass of greenhouses. The walls of glass act as a thermal trap for radiant heat, and so greenhouses tend to be warmer inside than the air outside (see also Section 12.10).

Microwaves and radar waves have longer wavelengths than infra-red light. Microwaves are used in some electric ovens for cooking. They are sufficiently penetrating to cook meat all through at the same time, instead of from the surface inwards. Radar waves are strongly reflected by concrete and metal structures, and they are widely used for navigation and communication purposes.

Radio and TV waves have greater wavelengths than all the other radiations in this extended electromagnetic spectrum. They are generated by oscillating electric charges in aerials, and received also by means of aerials. The longest radio wavelengths regularly in use are of about 2000 m.

Table 19.1 includes details of the radiations which make up the electromagnetic spectrum, listed in order of increasing wavelength. Typical wavelengths in air are given, although the radiations have various wavelengths which together form a continuous range from the shortest (γ -rays) to the longest (radio waves). Electromagnetic waves have certain common properties. These include:

- (i) they all consist of oscillating electric and magnetic fields,
- (ii) they are all transverse waves (see Section 14.3),
- (iii) they can all travel in a vacuum,
- (iv) in a vacuum or free space, they all travel at the speed of light, which is 3×10^8 m/s.

The frequencies of electromagnetic waves are found using the wave equation (14.1) which applies to all wave motions. Since $v = f\lambda = 3 \times 10^8$ m/s

Table 19.1 *the electromagnetic spectrum of radiation*

<i>Wave names</i>	<i>Typical wavelengths (m)</i>	<i>Emitters</i>	<i>Detectors</i>
γ -rays	10^{-14}	Radioactive nuclei	GM tube and photograph
X-rays	10^{-10}	X-ray tube	Photograph
Ultra-violet rays	10^{-8}	Quartz lamp, Sun, mercury vapour lamp	Photograph, fluorescent paper, photo-electric cell
Visible light	5×10^{-7}	Sun, filament lamp, strip lighting	Photograph, eye phototransistor, photoelectric cell
Infra-red rays	10^{-5}	Sun, electric fires, hot solids	Phototransistor, blackened bulb thermometer, skin
Microwaves	10^{-2}	Some electric ovens	Electronic detector
Radar waves	3×10^{-2} 1 300	Oscillators and transmitter aerials	Aerials, cathode ray tubes, TV sets, radio telescope, aerial and receiver
VHF and TV transmissions			
Radio waves			

for all electromagnetic waves in free space, it follows that the shortest wavelengths are emitted by sources which have the highest frequencies (γ -rays). At the other end of the spectrum, radio waves have the longest wavelengths and the lowest frequencies.

19.4 COLOUR ADDITION

If a white screen is illuminated simultaneously by two different coloured lights, a third colour appears on the screen where the original colours overlap. For example, yellow is seen when red and green light beams illuminate the screen and overlap. Experiments show that

Red + Green = Yellow

Green + Blue = Peacock blue (called cyan or turquoise)

Blue + Red = Magenta

The colours red, green and blue cannot be produced by adding other colours on the screen; they are called *primary colours*. When two primary colours are added, the resulting colours are called *secondary colours*. Yellow, peacock blue and magenta are therefore secondary colours.

When all three primary colours are added, the screen appears white, i.e.

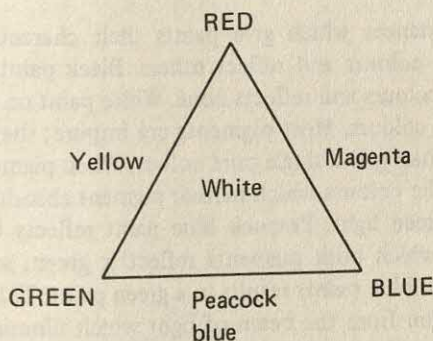
Red + Blue + Green = White

The screen also appears white when the three secondary colours are added, i.e.

Peacock blue + Magenta + Yellow = White

The above results can be expressed by means of a colour triangle which has the primary colours at its vertices, secondary colours at the midpoints of sides, and white in the middle (see Fig. 19.5). The figure is obtained by putting around the triangle what are almost the rainbow colours taken in sequence.

fig 19.5 the colour triangle



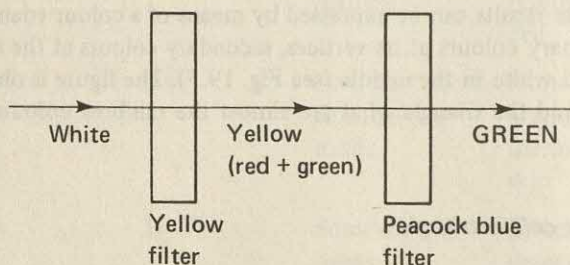
When white light is produced by the addition of two colours, they are described as *complementary colours*. Complementary colours consist of a primary colour and the secondary colour at the opposite side of the triangle. For example, green and magenta are complementary colours; when they are added, they produce white light.

19.5 COLOUR SUBTRACTION

(a) Colour filters

A colour filter is described by the colour of light which passes through it. A red filter, for example, passes red light and absorbs light of other colours. A yellow filter passes yellow light but also some red and green light which are next to yellow in the spectrum. A peacock blue filter allows green and blue light to pass through it; and a magenta filter allows red and blue light to pass while absorbing the other colours in the spectrum. Fig. 19.6 shows a primary colour being produced from a beam of white light by means of two secondary filters. In this example, green is the only colour which passes through both the yellow and the peacock blue filters.

fig 19.6 *white light is passed through a yellow and a peacock blue filter, and green light emerges*



(b) Pigments

Pigments are substances which give paints their characteristic colours. They absorb some colours and reflect others. Black paint has a pigment which absorbs all colours and reflects none. White paint on the other hand reflects light of all colours. Most pigments are impure; they reflect bands of colours rather than just a single pure colour. When pigments are mixed, they reflect only the colours which neither pigment absorbs. Yellow paint reflects red and green light. Peacock blue paint reflects blue and green light. The colour which both pigments reflect is green, so a mixture of yellow and peacock blue paints results in a green paint. This is an example of colour subtraction from the beam of light which illuminates the paint. It is comparable with the example for colour filters given in Fig. 19.6.

19.6 COLOURED OBJECTS SEEN IN COLOURED LIGHTS

The apparent colour of an object depends on the colour of light illuminating it. A yellow object reflects red and green light and absorbs blue.

Hence, a yellow object illuminated by red light appears red, because red light is the only colour reflected. If a yellow object is illuminated by blue light, it appears black, because the blue light is absorbed and no light is reflected.

A red object appears red when seen in yellow or magenta light. Both colours contain red which is the only colour that the red object reflects. But it appears black if illuminated by peacock blue light, because peacock blue contains no red light; the colours in peacock blue are green and blue, both of which are absorbed by a red object.



SOUND

20.1 INTRODUCTION

All sources of sound have some part of them which is vibrating, for example, a violin string or the surface of a drum. Sound travels in the form of longitudinal waves, that is, molecules vibrate to and fro in the direction of travel of the sound (see Section 14.4). When sound is heard, energy is carried from the source of the waves to the ear of the listener; but the molecules of air in between do not move as a whole towards the listener (see Section 14.1). At any instant there are regions where the air is compressed (*compressions*), separated by regions where the air is rarefied (*rarefactions* or *decompressions*). Sound waves consist of a series of alternate compressions and rarefactions travelling away from a source at a certain speed determined by the nature of the medium in which they flow.

Sound cannot be transmitted through a space where there is no medium. An electric bell inside a bell jar cannot be heard after the air has been pumped out, but its loudness increases as air is allowed back into the jar. In the same way, sounds cannot be transmitted above the Moon's surface where there is no atmosphere; the Moon is sometimes called the 'silent planet'.

Sound cannot travel through a vacuum

The wavelength of sound waves is the distance between two adjacent compressions or two adjacent rarefactions (Section 14.2)

The frequency of sound waves is the number of waves passing a point every second. The unit of frequency is the hertz (Hz), as it is for other wave motions. $1 \text{ Hz} = 1 \text{ cycle per second}$, which means that the molecules vibrate backwards and forwards once every second

Like all other wave motions, sound waves can be reflected; echoes from walls obey the laws of reflection (Section 15.2). They can also be refracted. A balloon filled with carbon dioxide acts like a lens and focuses sound waves.

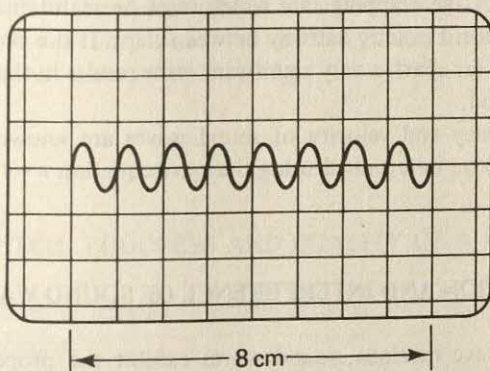
20.2 MEASUREMENT OF THE FREQUENCY OF SOUND WAVES

(a) Using a cathode ray oscilloscope

A microphone converts sound energy into electrical energy. It responds to the small pressure variations in sound waves and converts them into small voltage variations of the same frequency. When a microphone is connected via an amplifier to the input terminals of a cathode ray oscilloscope, the waveform of sound reaching the microphone can be displayed on the screen of the cathode ray tube.

Suppose the time base is set at 1 millisecond per centimetre (1 ms/cm), and that the trace on the face of the oscilloscope looks like the one shown in Fig. 20.1. Then one complete sound wave reaches the microphone as

fig 20.1



the trace sweeps horizontally across 1 cm. This takes 1 ms, and if one wave arrives every millisecond, then there are 1000 waves/s. The frequency is therefore 1000 Hz.

(b) Using a stroboscope

In this method the source is 'frozen' by using a flashing light (see Section 14.5(d)). The maximum flashing rate when the source appears stationary is the frequency of the source, and of the waves it emits.

20.3 MEASURING THE VELOCITY OF SOUND IN AIR BY AN ECHO METHOD

The experiment is carried out in an open space some 50 to 100 m from a large vertical wall. A person stands a measured distance from the wall clap-

ping loudly at regular intervals. With a little practice the rate of clapping can be adjusted so that the echo returns from the wall midway between two claps. Only then does a second observer measure the rate of clapping using a stopwatch.

Suppose the timer standing 55 m from the wall counts 30 claps in 20 s. The time between claps is $20/30$ s, and the time between a clap and its echo is half as much, i.e. $20/60$ or $1/3$ s. In $1/3$ s sound travels to the wall and back, which is 110 m. The speed of sound in air is therefore

$$\frac{110}{1/3} = 330 \text{ m/s.}$$

Results obtained by this method vary considerably, suggesting that any one value may not be very accurate. Why is the result liable to be inaccurate? A small error would be introduced if the estimate of the distance to the wall were in error by a few centimetres; and the timing of 30 claps with a good stopwatch would introduce a small error too. By far the greatest source of error is the clapping rate which must be maintained steady so that echoes are heard exactly halfway between claps. If the claps are a few milliseconds too far apart, a very significant error results in the estimate of the speed of sound.

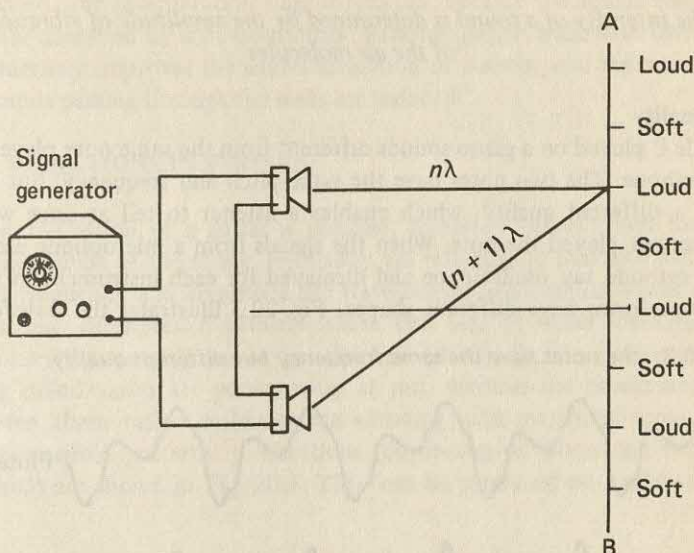
If the frequency and velocity of sound waves are known, the wavelength λ can also be determined using the wave equation $v = f\lambda$ (equation (14.1)).

20.4 DIFFRACTION AND INTERFERENCE OF SOUND WAVES

Like all other wave motions, sound waves exhibit the properties of diffraction and interference (see Section 14.9). Diffraction enables sound waves to be heard around the corners of buildings or over the top of a high wall. Sound can also be diffracted on passing through an open doorway whose width is of the order of the wavelength of the sound.

The interference of sound waves can be demonstrated using two loudspeakers connected to the same signal generator and spaced about 3 m apart, as illustrated in Fig. 20.2. Anyone walking along a line AB parallel to the line joining the loudspeakers hears a note of varying intensity or loudness. A soft note of minimum intensity is heard wherever a compression from one loudspeaker coincides with a rarefaction from the other loudspeaker; and a loud note of maximum intensity is heard wherever two compressions coincide followed a moment later by two decompressions from the two loudspeakers. The spacing of the loudest points along AB in Fig. 20.2 can be used to find the wavelength of the sound. The theory is given in Section 14.9(d) for the measurement of the wavelength of light using Young's slits and it applies to this situation as well as to light.

fig 20.2 the sound waves from the two sources interfere constructively if the path difference is a whole number of wavelengths



20.5 THE PITCH, LOUDNESS AND QUALITY OF A MUSICAL NOTE

A source vibrating irregularly produces noise, but regular vibrations produce *musical notes*.

(a) Pitch

The term *pitch* is used by musicians to distinguish different musical notes. The pitch of a note depends on its frequency; if the frequency increases, the pitch of the note goes up. Musical notes in the audible range have frequencies from about 20 Hz up to about 20 kHz; the range is greater for a dog's ears, and dog whistles usually have frequencies which are higher than human ears can hear.

(b) Loudness

The term *loudness* is used to describe our response to sound reaching our ears. A loud sound is more easily heard than a soft sound. But loudness is not a very precise description; and *intensity* is often used by physicists as a measure of the sound energy reaching the ear every second.

The intensity of sound at a particular place is the power (energy per second) transmitted through unit area at that place

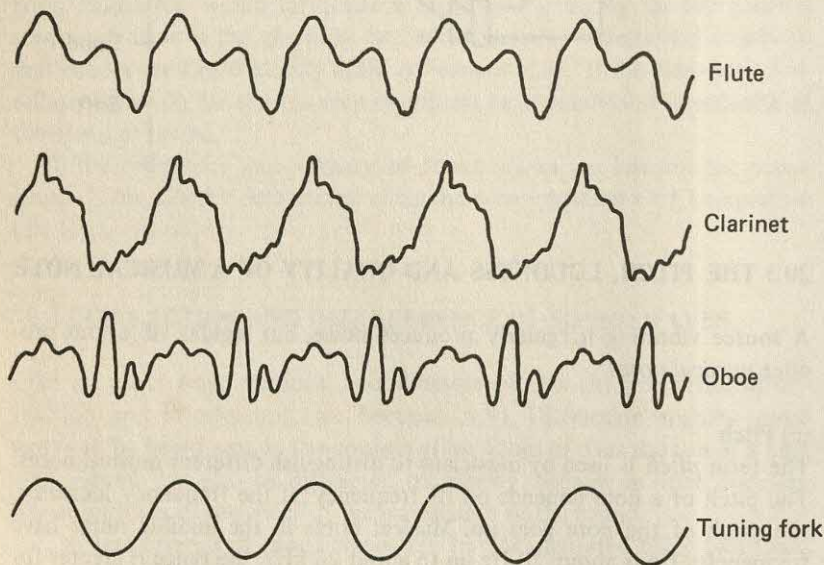
The intensity of the sound emitted by a source depends on the amplitude of vibration of the source (see Section 14.8). The intensity increases with the amplitude of vibration.

The intensity of a sound is determined by the amplitude of vibration of the air molecules

(c) Quality

Middle C played on a piano sounds different from the same note played on a trombone. The two notes have the same pitch and frequency; but they have a different quality, which enables a listener to tell at once which instrument played the note. When the signals from a microphone are fed to a cathode ray oscilloscope and displayed for each instrument in turn, the waveforms have different shapes. Fig. 20.3 illustrates the waveforms

fig 20.3 *the notes have the same frequency but different quality*



of a note played on a flute, clarinet and oboe. In each waveform the quality is different from that of a tuning fork of the same frequency. The pure note of the tuning fork has the simplest waveform.

20.6 ABSORPTION OF SOUND WAVES

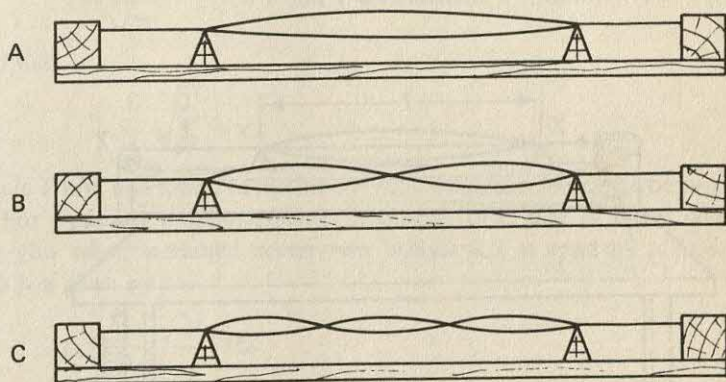
Sound waves are normally partially absorbed and partially reflected by materials in their path. A brick wall is a good reflector of sound, and a

cushion is a good absorber. If a sound is made in an empty room having no furniture or curtains, the sound will *reverberate*. It will die away gradually as it is repeatedly reflected to and fro across the room. When the room is furnished, the reverberation time is reduced, because the sound energy is quickly absorbed by soft furnishings. Packing hollow walls with fibreglass considerably improves the sound insulation of a room, and the intensities of sounds passing through the walls are reduced.

20.7 VIBRATIONS IN STRINGS

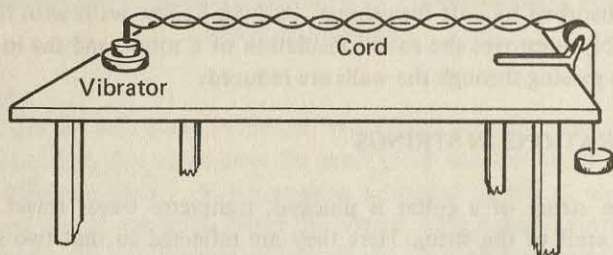
When the string of a guitar is plucked, transverse waves travel towards the fixed ends of the string. Here they are reflected so that two waves of equal frequency and amplitude are travelling in opposite directions along the string. Interference between these two sets of waves gives rise to *stationary* or *standing waves* (see Section 14.10). Certain points on the string called *nodes* are permanently at rest, whereas the points midway between them called *antinodes* are vibrating with maximum amplitude. Three possible patterns of vibrations (depending on where the string is plucked) are shown in Fig. 20.4. They can be produced on a rubber cord

fig 20.4 *three possible modes of vibration of a string*



using an electrical vibrator and studied in stroboscopic light (Fig. 20.5). Pattern A (Fig. 20.4) shows the note of lowest pitch, called the *fundamental*. In pattern B the wavelength has been halved and the frequency doubled. This results in a note called the first *overtone* which is one octave above the fundamental. Pattern C is the second overtone and is produced when the string vibrates at three times the frequency of the fundamental.

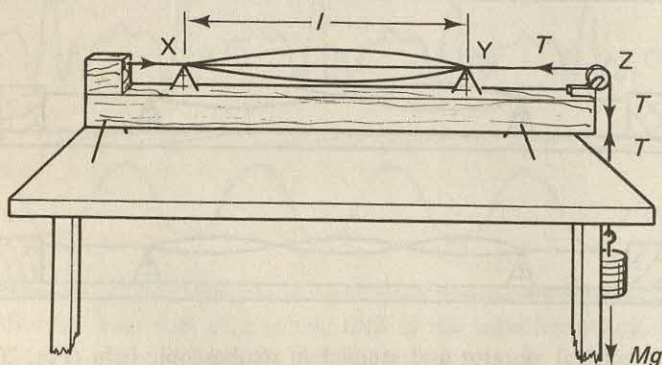
fig 20.5 *the vibrator moves up and down. Transverse waves move down the rubber cord and are reflected at the far end producing stationary waves in the rubber cord*



20.8 FREQUENCY OF A VIBRATING STRING

A sonometer (Fig. 20.6) may be used to investigate the various factors which determine the natural frequency of vibration of a fixed string. It has a wooden base with a wire fixed to a pin at one end. The wire passes over two bridges X and Y, and finally over a pulley Z. Standard masses are hung on the free end of the wire. The note emitted when the wire is

fig 20.6 *a sonometer*



plucked midway between X and Y has a frequency which varies with the length of wire l between the bridges. When the pitch of its note matches that of a tuning fork or oscillator, the frequency f and the length l are recorded. This is repeated using a series of tuning forks, and a series of readings of f and l is obtained using the same mass hanging at the end of the wire (the

tension in the wire therefore remains constant). It is found that doubling the length of wire halves its natural frequency of vibration; so

$$f \propto 1/l$$

This means that f is inversely proportional to l .

When the apparatus is used to investigate how f varies with the tension T in the wire, the length l is kept constant. The loads hung at the end of the wire are varied until its note matches that of the standard tuning fork; then readings of the tension in the wire supporting the load and the frequency of the note are recorded. A series of readings is obtained using different tuning forks.

A third series of experiments requires wires of the same length and tension, but different mass m per unit length of wire. Various wires having different linear densities are chosen and the frequencies of their emitted notes recorded.

Analysis of the results of the above experiments show that for the fundamental note

$$f \propto \frac{1}{l} \quad \text{if } T \text{ and } m \text{ are constant}$$

$$f \propto \sqrt{T} \quad \text{if } l \text{ and } m \text{ are constant}$$

$$f \propto \frac{1}{\sqrt{m}} \quad \text{if } T \text{ and } l \text{ are constant}$$

and that

$$f = \frac{1}{2l} \sqrt{\frac{T}{m}}$$

where T is in newtons, l in metres and m in kilograms per metre of wire.

For example, the fundamental frequency of a wire of linear density 2.5 g/m when stretched across two bridges 0.3 m apart by a force of 100 N is given by :

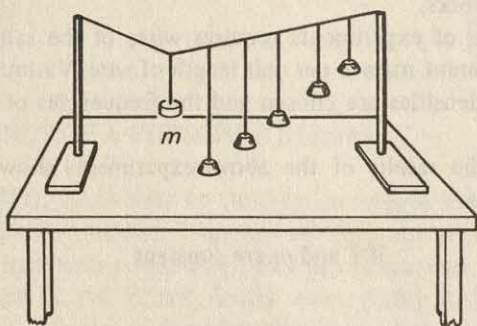
$$\begin{aligned} f &= \frac{1}{2l} \sqrt{\frac{T}{m}} \text{ Hz} \\ &= \frac{1}{2 \times 0.3} \sqrt{100 \times \frac{1000}{2.5}} \\ &= \frac{1}{0.6} \sqrt{40\,000} \\ &= \frac{200}{0.6} \end{aligned}$$

20.9 RESONANCE

All objects have their own natural frequencies of vibration. A pendulum when set in motion swings to and fro with its own natural frequency; and another pendulum of different length has a different natural frequency.

The experiment illustrated in Fig. 20.7 may be used to demonstrate the

fig 20.7 *Barton's pendulums. The light pendulum which has the same length of string as the mass m resonates with large amplitude when the mass m swings to and fro*



phenomenon of *resonance*. A series of light pendulums of different lengths is supported from a single string which also carries a pendulum with a more massive bob. When the bob is drawn to the side and released, each of the light pendulums also swings with what are called *forced oscillations*. The amplitudes of the swings of the light pendulums are small except for the one whose natural frequency coincides with that of the massive pendulum. In this light pendulum, large amplitude oscillations build up. The phenomenon is called resonance.

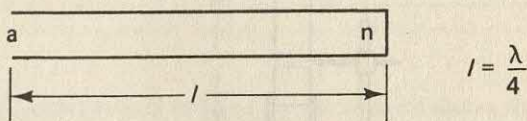
Where resonance occurs, the frequency of forcing oscillations is the same as the natural frequency of the system being forced to oscillate. At this frequency, oscillations of large amplitude result from applied impulses of small amplitude

20.10 STANDING WAVES AND RESONANCE IN AIR COLUMNS

A column of air in a pipe has its own natural frequency which varies with the length of the pipe. If you blow across the open end of a glass bottle, you may hear the note which the enclosed column of air naturally emits. The frequency of the note depends on the length of the air column in the bottle and does not depend on how hard you blow. A note of higher pitch is emitted if the bottle is partly filled with liquid, thus shortening the length

of the air column. The note results from standing waves set up in the air column, as pressure disturbances in the neck of the vessel travel to the sealed end and are reflected back along their path. The maximum amplitude of these disturbances occurs at the open end of the vessel or pipe, so the end of the pipe is an antinode of the displacements in the waves. At the closed end of a pipe, there can be no displacement of air and this is a node in the standing wave. Fig. 20.8 illustrates the positions of node and anti-

fig 20.8 when the fundamental is sounding, the length of the pipe is $\lambda/4$

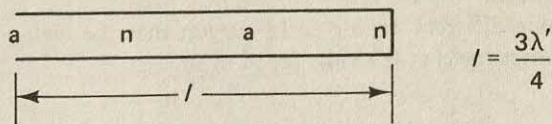


node in a pipe of length l . In the fundamental mode of vibration, l is the distance between a node and the nearest antinode. The distance between one node and the next in a standing wave is one-half a wavelength (see Section 14.10) so the length of a pipe from a node to an antinode is one-quarter a wavelength, i.e.

$$l = \lambda/4$$

Organ pipes sealed at one end are made in lengths chosen to give rise to fundamental notes of frequencies in a musical scale. But they can also emit overtones at the same time, giving different qualities to the sounds they emit. Fig. 20.9 illustrates the first overtone for the pipe of length l . From

fig 20.9 the first overtone. The length of the pipe is $3\lambda'/4$



the spacing of nodes and antinodes arranged so that one end of the pipe is at a node and the other end at an antinode, it appears that the length l is now three-quarters of a wavelength, i.e.

$$l = 3\lambda'/4$$

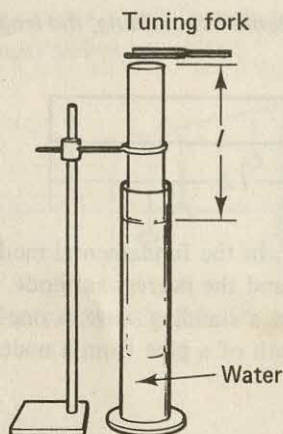
For the second overtone

$$l = 5\lambda'/4$$

Normally most of the sound energy is emitted at the frequencies of the fundamental and the first few overtones.

Resonance in an air column can be demonstrated using an open glass tube lowered slowly into a tall column of water while a tuning fork vibrates above the tube. As the tube is lowered into the position illustrated in Fig. 20.10, the column of air that it contains becomes shorter. When the

fig 20.10 *when the length of the air column is $\lambda/4$, the air in the tube resonates emitting a loud note*



natural frequency of the air column matches that of the struck tuning fork, the sound of the fork is heard more loudly in the room. The shortest air column that resonates with the tuning fork has a length equal to one-quarter of a wavelength of the sound in air. By measuring the length of the air column one can determine the wavelength; and as the frequency of the tuning fork is known, the velocity of sound in air can be calculated using the wave equation $v = f\lambda$ (equation (14.1)).

Results using different tuning forks suggest that the high-pitched and low-pitched notes travel at the same speed in air.

PART IV

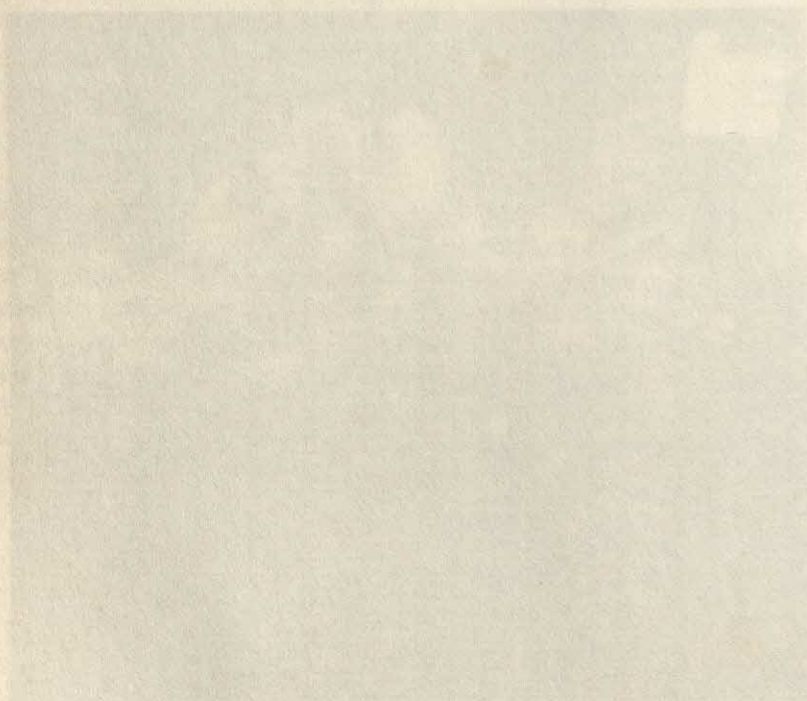
ELECTRICITY, ATOMIC AND

NUCLEAR PHYSICS

The lights of Piccadilly Circus in London (courtesy United Kingdom Atomic Energy Authority)



PART IV

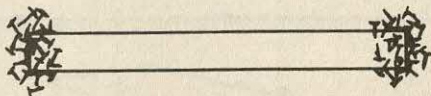
ELECTRICITY, ATOMIC AND
NUCLEAR PHYSICS

MAGNETISM

21.1 PROPERTIES OF MAGNETS

If you have played with magnets or magnetic toys, you will probably be familiar with two of their basic properties, namely the fact that they attract pieces of iron or steel (Fig. 21.1) and, secondly, if pivoted or sus-

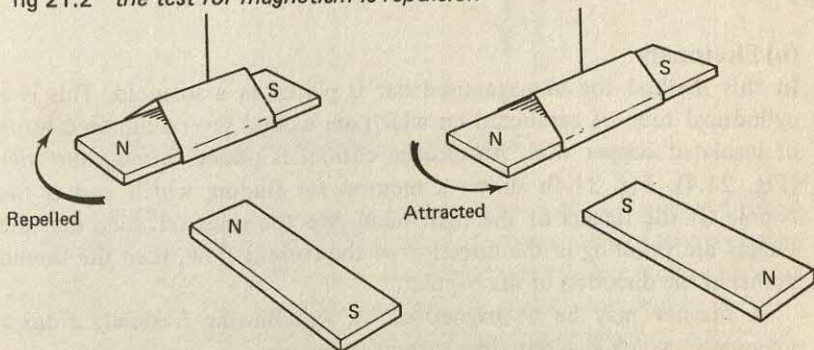
fig 21.1 *the nails adhere to the regions at the ends of the magnet known as the poles*



pended, they always come to rest pointing in a definite direction. When pivoted the end which points towards the North of the Earth is called the *North-seeking pole* or simply the N-pole. The end which points South is called the *South-seeking pole* or the S-pole. (Two substances other than iron and steel which are attracted by magnets are cobalt and nickel.)

If the N-pole of a magnet is brought near the N-pole of another magnet, then repulsion occurs (Fig. 21.2). Two S-poles will also repel each other

fig 21.2 *the test for magnetism is repulsion*



while a N-pole and a S-pole will attract each other. We may summarise this by saying:

Like poles repel, unlike poles attract

21.2 THE TEST FOR MAGNETISM

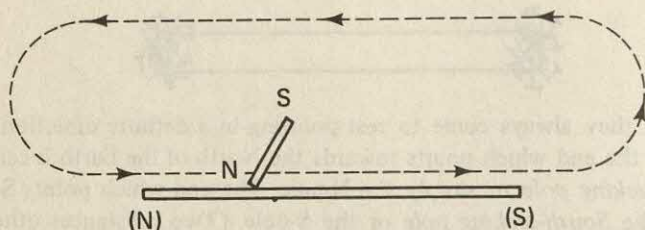
If you want to discover whether a bar is magnetised or not, bring each end in turn up to the N-pole of a suspended magnet (Fig. 21.2). If one end repels the N-pole, then the bar is magnetised. Repulsion is the only sure test for a magnet. An unmagnetised bar will attract the N- and the S-pole of a suspended magnet.

21.3 MAKING A MAGNET

(a) By stroking

Fig. 21.3 shows one method of magnetising a steel knitting needle. The needle is stroked in the same direction a number of times. After each

fig 21.3 making a magnet



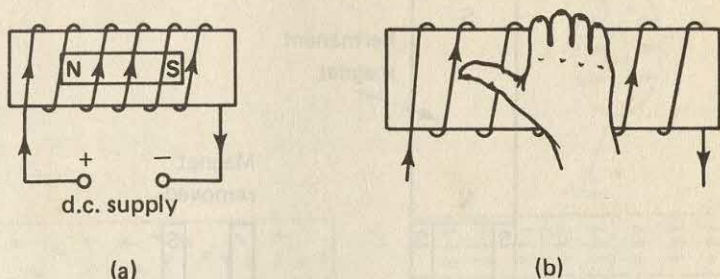
stroke the magnet must be lifted well clear of the needle. The end last touched by the magnet has the opposite polarity to that of the stroking pole.

(b) Electrically

In this method the unmagnetised bar is placed in a solenoid. This is a cylindrical tube of cardboard on which are wound several hundred turns of insulated copper wire. An electric current is passed through the wire (Fig. 21.4). Fig. 21.4b shows a method for finding which end is the N-pole. If the fingers of the right hand grip the solenoid, such that the fingers are pointing in the direction of the current flow, then the thumb points in the direction of the N-pole.

A magnet may be de-magnetised by withdrawing it slowly from a solenoid in which an alternating current is flowing.

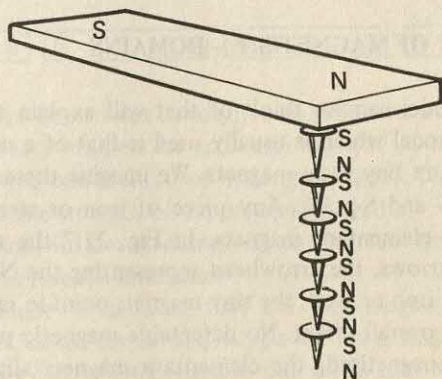
fig 21.4 magnetising a bar. In (b) the fingers are pointed in the direction of the current; the N-pole is at the end to which the thumb is pointing



21.4 INDUCED MAGNETISM

It is very easy to pick up a whole chain of nails using a bar magnet (Fig. 21.5). The nails become magnetised as shown in the diagram. They are said to be magnetised by *induction*. If the magnet is removed the nails cease to be magnetised.

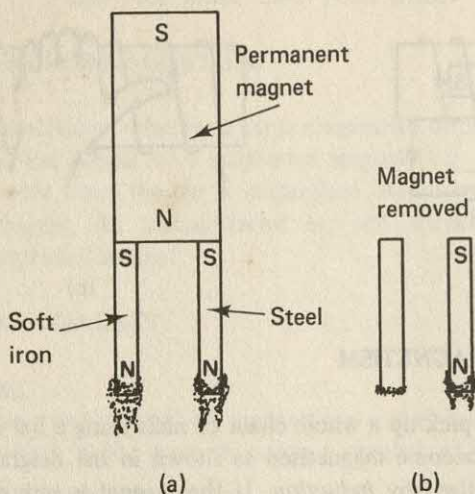
fig 21.5 the tacks become magnets by induction



21.5 PROPERTIES OF SOFT IRON AND STEEL

Fig. 21.6a shows magnetism being induced in bars of iron and steel. When the magnet is removed (Fig. 21.6b) the steel retains most of its magnetism but the iron loses its magnetism. Magnetic materials like iron, which are easy to magnetise but do not retain their magnetism, are said to be 'soft'. Those like steel, which are harder to magnetise but retain their magnetism,

fig 21.6 *soft iron loses its magnetism when the permanent magnet is removed*



are said to be 'hard'. Permanent magnets are made from materials like steel. Electromagnets (see Section 24.3) are made from soft iron.

21.6 A THEORY OF MAGNETISM – DOMAINS

What sort of model can we think of that will explain the properties of magnets? The model which is usually used is that of a magnetic material consisting of many tiny little magnets. We imagine these tiny elementary magnets have N- and S-poles. Any piece of iron or steel contains many millions of such elementary magnets. In Fig. 21.7 the tiny magnets are represented by arrows, the arrowhead representing the N-pole. In a piece of unmagnetised iron or steel, the tiny magnets point in random directions and produce no overall effect. No detectable magnetic poles exist. When the material is magnetised, the elementary magnets align themselves as shown in the diagram. The N- and S-poles in the middle annul one another (see Fig. 21.7b), but one end has N-poles un-neutralised and the other end has S-poles un-neutralised. One end therefore behaves as a N-pole and the other as a S-pole. The properties mentioned above can easily be explained using this model. Can we test the model still further?

One way to test it is to magnetise a watch spring as shown in Fig. 21.8 and then break it into a number of pieces. Each piece has N- and S-poles as shown in the diagram, and this would clearly be expected from our theory of tiny elementary magnets. The theory also explains why a magnet

fig 21.7 when a magnetic substance is fully magnetised, all the tiny elementary magnets point in one direction

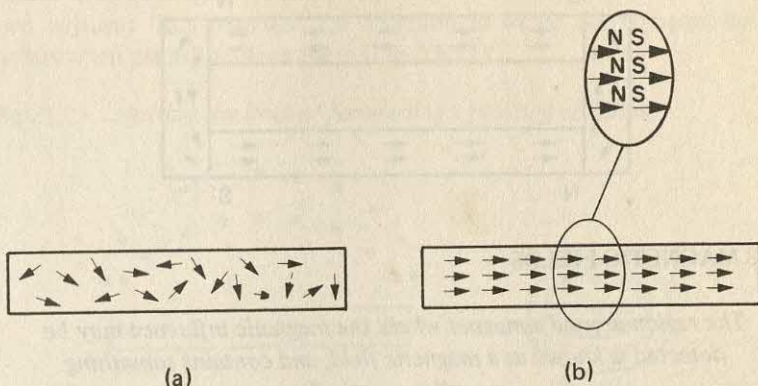
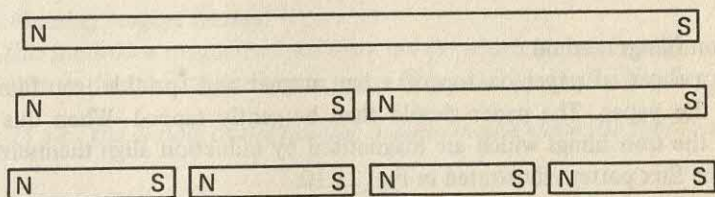


fig 21.8 when a magnet is broken, new poles are created



heated in a bunsen flame loses its magnetism. The atoms vibrate and jostle about so that the tiny magnets are no longer aligned.

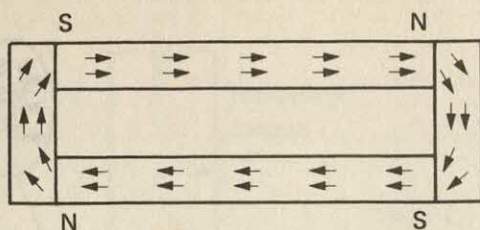
Fig. 21.3 showed one method of making a magnet. Continued stroking does not increase the strength of the magnet indefinitely. Once all the tiny magnets are aligned, further stroking will not increase the magnetism. We say the magnet has reached *saturation*.

There is evidence to indicate that the tiny magnets are groups of millions of atoms, called *domains*. Can you use the theory of tiny magnets to show how a ring can be magnetised so that it has got no poles?

21.7 KEEPERS

A magnet tends gradually to lose its magnetism because the like poles of the tiny magnets at each end tend to repel each other. To prevent this loss of magnetism, bar magnets are often stored in pairs with unlike poles at opposite ends with pieces of soft iron, called *keepers*, across the ends (Fig. 21.9). The induced magnetism in the keeper means that there are no longer any 'free' poles and the tiny magnets tend to remain aligned.

fig 21.9 keepers



21.8 MAGNETIC FIELDS

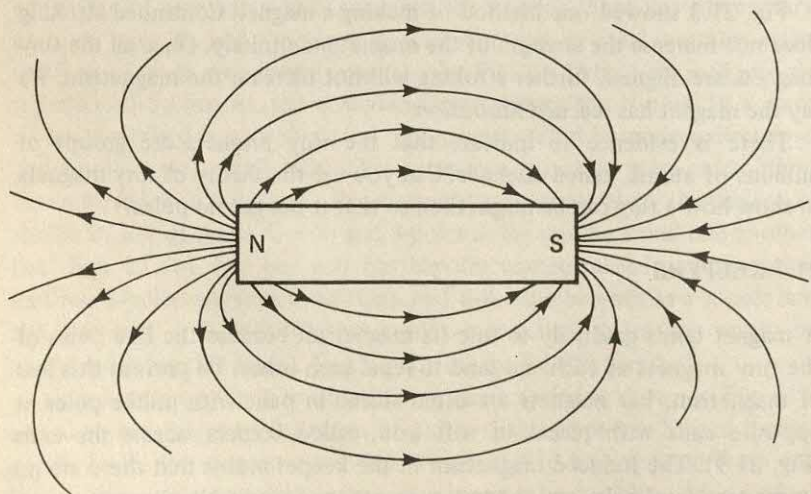
The region around a magnet where the magnetic influence may be detected is known as a magnetic field, and contains something we call magnetic flux

There are several ways of showing the presence of this invisible magnetic flux.

(a) Iron filings method

Place a sheet of paper on top of a bar magnet and sprinkle iron filings onto the paper. The paper should then be gently tapped. When this is done, the iron filings which are magnetised by induction align themselves into the flux pattern illustrated in Fig. 21.10.

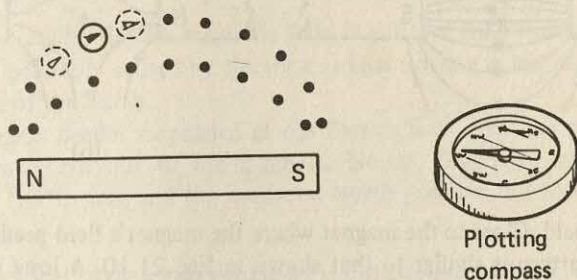
fig 21.10 the flux patterns or lines of force around a bar magnet



(b) Plotting compass method

The magnet is placed on a sheet of paper and the plotting compass is placed near the magnet. A series of dots are put on the paper such that any two adjacent dots represent the direction in which the compass needle points when placed between them (Fig. 21.11).

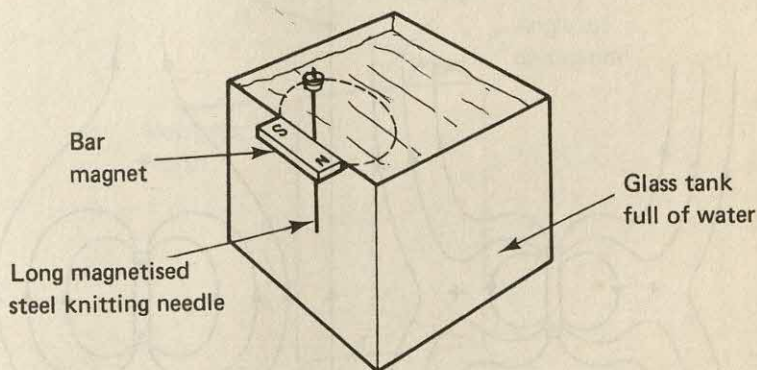
fig 21.11 *plotting the lines of force using a plotting compass*



(c) Floating magnet method

In this method a magnetised knitting needle which passes through a piece of cork is floated vertically on water (Fig. 21.12). If the N-pole of the

fig 21.12 *the N-pole of the floating magnet moves along the flux line*

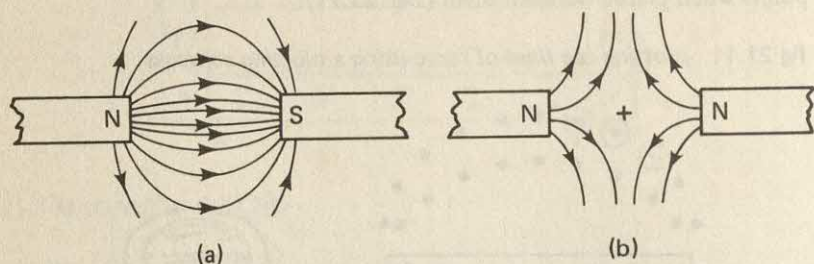


vertical needle is positioned near the N-pole of the bar magnet and then released, it will move along the *line of force* as indicated in the diagram.

A magnetic line of force shows the direction of the field at any point. It is the direction in which a free N-pole would move if placed in the field

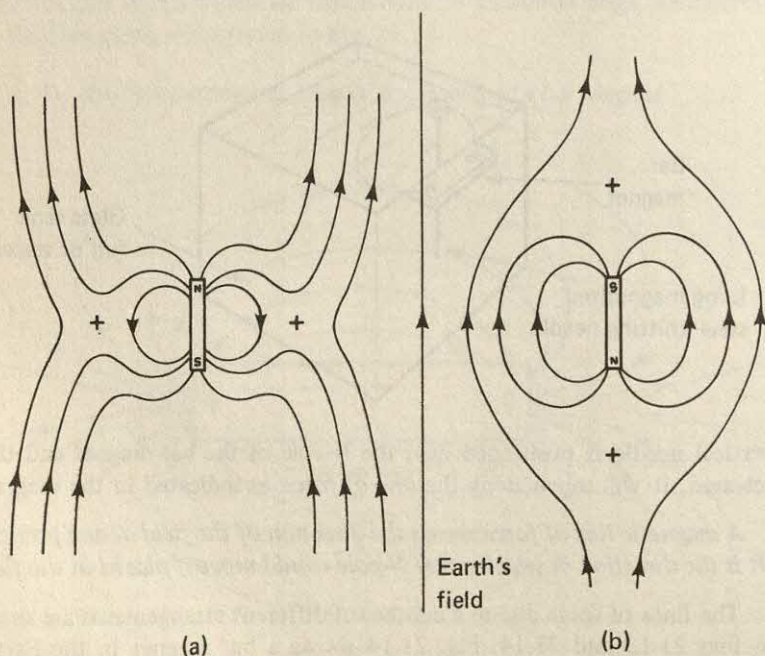
The lines of force due to a number of different arrangements are shown in Figs 21.13 and 21.14. Fig. 21.14 shows a bar magnet in the Earth's

fig 21.13 *the flux patterns resulting from two bar magnets: (a) with opposite poles facing each other and (b) with like poles facing each other*



magnetic field. Close to the magnet where the magnet's field predominates, the flux pattern is similar to that shown in Fig. 21.10. A long way away from the magnet where the Earth's magnetic field is stronger, the flux patterns are shown by parallel lines pointing magnetic North and South.

fig 21.14 *flux patterns resulting from a magnet placed in the Earth's magnetic field*



The crosses show the *neutral points*. Neutral points are where the fields due to the magnet and the Earth are exactly equal and opposite. The resultant field at such a point is zero.

Notice that the lines of force representing the flux patterns can never touch or cross.

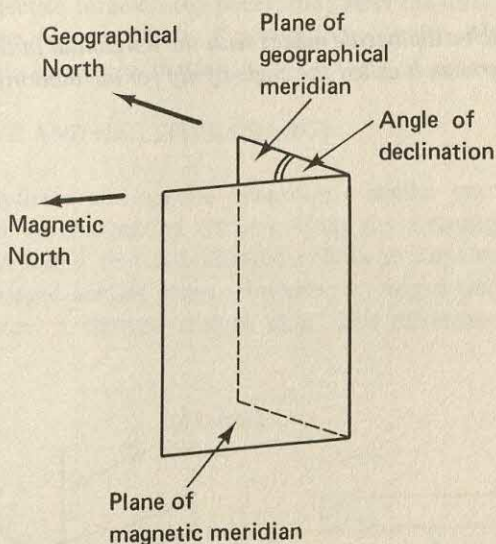
21.9 SOME CHARACTERISTICS OF THE EARTH'S MAGNETIC FIELD

The origin of the Earth's magnetic field is still not fully understood, but it is most probably caused by electric currents flowing in the liquid core at the centre of the Earth.

A compass needle suspended in the Earth's field does not point to the geographic North but to the magnetic North. That is to say, the geographical North pole and the magnetic North pole are not in exactly the same place.

The vertical plane which contains the direction in which a compass needle sets when suspended in the Earth's field is called the magnetic meridian at the point (Fig. 21.15)

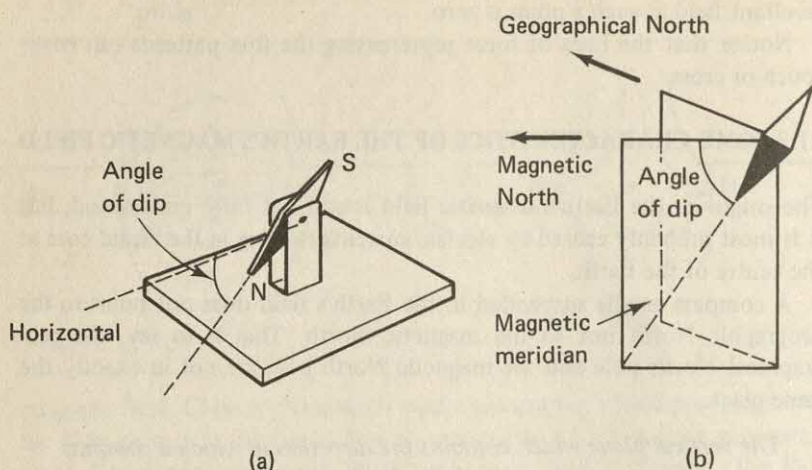
fig 21.15



The angle between the magnetic and geographic meridians is called the declination (or variation)

A magnet which is pivoted at its centre of gravity so that it is able to rotate in a *vertical* plane is called a dip needle (Fig. 21.16a). When such a

fig 21.16 (a) a dip needle. (b) The dip needle is pivoted so that it can rotate in the magnetic meridian



magnet is placed in the magnetic meridian in the northern hemisphere, the North pole points downwards. The angle the magnet makes with the horizontal (in the UK) is about 65° .

The angle which a dip needle makes with the horizontal in the magnetic meridian is called the angle of dip (or inclination)

ELECTRONS AND ELECTRON BEAMS

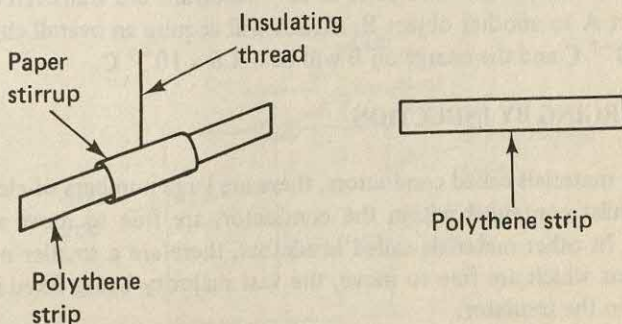
22.1 CHARGING BY FRICTION

When a plastic ruler is placed near some very tiny pieces of paper, it will not have any effect on them. However, if the ruler is rubbed with a duster, it is found that the pieces of paper are attracted to the ruler. This difference in behaviour is explained by saying that the ruler has become electrically charged when rubbed with the duster. Because the ruler is charged, it exerts an electric force on the paper; the nearer the ruler is to the paper the greater the force it exerts. Charging the ruler by rubbing it with a duster is called *charging by friction*.

22.2 POSITIVE AND NEGATIVE CHARGE

Strips of polythene and acetate behave in a similar way to the plastic ruler and can be charged by friction. Using the arrangement shown in Fig. 22.1 it is found that two charged polythene strips repel each other as do two charged acetate strips. However, a charged polythene strip is found to attract a charged acetate strip. This difference in behaviour

fig 22.1



indicates that the charge on the acetate is different from that on the polythene strip. To account for this difference, the acetate strip is said to be positively charged and the polythene is said to be negatively charged.

All charged objects are found to have a charge similar to that of either the acetate or polythene, that is, to be positively or negatively charged. It is always found that objects with the same charge repel each other, whereas objects with opposite charge attract each other.

Like charges repel and unlike charges attract

We can now explain why a piece of paper is attracted by a charged rod. Suppose a negatively charged rod is held above a small piece of paper. The electrons in the paper will be repelled leaving the top of the paper positively charged. The attraction of unlike charges (negative on the rod and positive on top of the paper) results in the paper being attracted by the rod.

22.3 PROTONS AND ELECTRONS

The existence of two and only two types of charge comes about from the nature of the atom itself.

All matter is composed of atoms which consist of particles called protons, neutrons and electrons. Protons have a positive charge, neutrons have no charge and electrons have a negative charge. In an atom the number of protons equals the number of electrons, and as the charge on the proton is the same size as that on the electron, the atom is electrically neutral, that is, the amounts of positive and negative charge are the same.

If some electrons are removed from an object it will acquire an overall positive charge, as the amount of positive charge will now be greater than the amount of negative charge. This is exactly what occurs when an acetate strip is rubbed with a duster. Some of the electrons in the acetate are transferred from the strip to the duster, the strip becoming positively charged, and the duster becoming negatively charged.

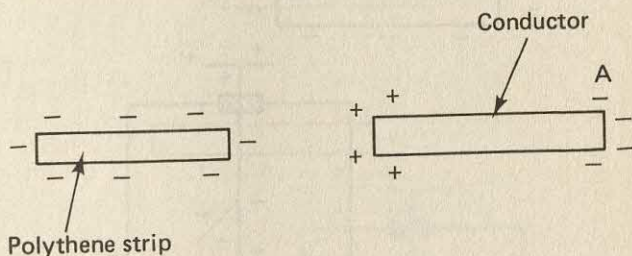
Electric charge is measured in coulombs (C), and the charge on one electron is -1.6×10^{-19} C. Thus if 10^{12} electrons are transferred from one object A to another object B, then A will acquire an overall charge of $+1.6 \times 10^{-7}$ C and the charge on B will be -1.6×10^{-7} C.

22.4 CHARGING BY INDUCTION

In certain materials called conductors, there are large numbers of electrons which, whilst contained within the conductor, are free to move around within it. In other materials called insulators, there are a smaller number of electrons which are free to move, the vast majority being fixed in position within the insulator.

As they contain a large number of free electrons, conductors can be charged by a process known as *induction*. In order to do this, a charged object such as a polythene strip is placed near the conductor (Fig. 22.2). The negative charge on the polythene repels some of the free electrons in the conductor to A. If you touch the conductor at A, some of the electrons will move from the conductor to the earth. If you now remove your finger and then remove the polythene strip, the conductor, having *lost* some electrons, will be positively charged.

fig 22.2 *charging by induction*



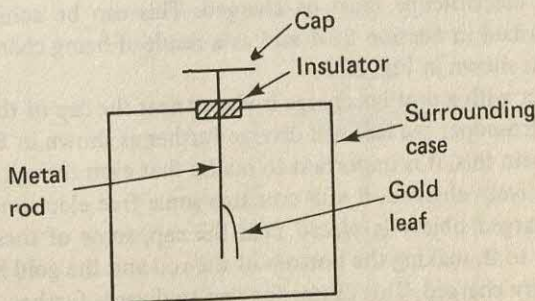
Similarly if a positively charged acetate strip is brought near a conductor and the same procedure adopted, the conductor will become negatively charged. In this case electrons are transferred from the earth (via your finger) *to* the conductor, making it negatively charged.

It is important to note that when charging by induction the conductor obtains a charge *opposite* to that of the object used in charging it.

22.5 THE GOLD-LEAF ELECTROSCOPE

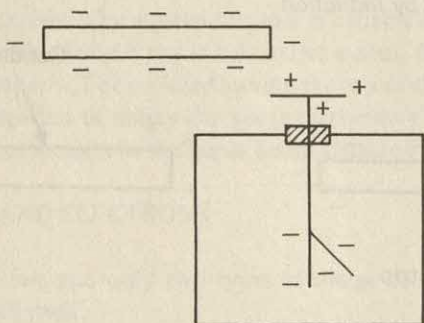
This is a convenient instrument for detecting whether or not an object is charged. As shown in Fig. 22.3, it consists of a small piece of gold foil

fig 22.3 *a gold-leaf electroscope. Notice that the cap and rod are insulated from the case*



connected to the lower end of a metal rod. The object to be tested is placed near the cap as shown in Fig. 22.4. If the object is negatively charged, some of the electrons in the cap will be repelled onto the bottom end of the rod and also onto the gold foil. As the gold foil and the bottom of the rod are now *both* negatively charged, the foil will be deflected as shown in Fig. 22.4.

fig 22.4 testing whether an object is charged



Similarly, if the object is positively charged, the gold leaf will behave in the same way. In this case, the bottom end of the rod and the gold leaf will be positively charged, as some of the electrons in the bottom of the rod and the gold leaf will be attracted to the cap.

If the object is uncharged, there will be no movement of electrons and the leaf will remain in the position shown in Fig. 22.3.

The procedure described above can detect whether or not the object is charged but it cannot be used to find the sign of the charge, as the gold leaf behaves in the same way whether the object is positively or negatively charged.

However, the instrument can be used to determine the sign of the charge on an object if the following procedure is adopted.

First, the electroscope must be charged. This can be achieved in the manner described in Section 22.4 and as a result of being charged the leaf will diverge as shown in Fig. 22.5.

If an object with a positive charge is placed near the cap of the positively charged electroscope, the leaf will diverge further as shown in Fig. 22.6. In order to explain this, it is important to realise that even though the electroscope is positively charged, it still contains some free electrons. When the positively charged object is placed near the cap, some of these electrons are attracted to it, making the bottom of the rod and the gold leaf become *more positively* charged. This causes the leaf to diverge further.

fig 22.5 a charged electroscope

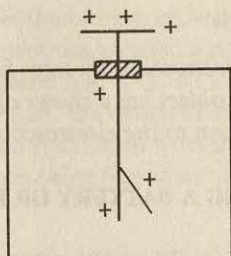
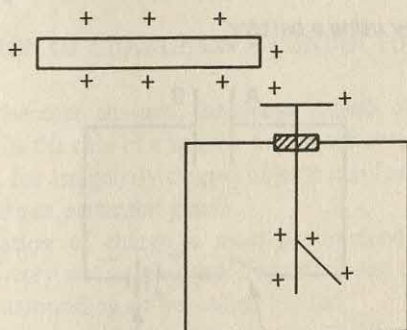


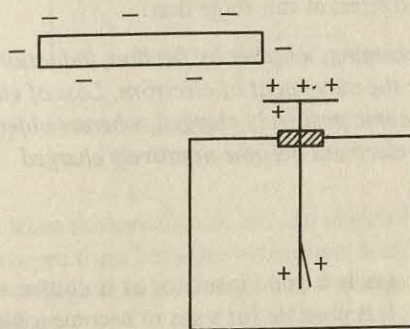
fig 22.6



If a negatively charged object is placed near the cap, the leaf will collapse as shown in Fig. 22.7. This occurs because electrons are repelled onto the bottom end of the rod and the gold leaf, making them *less positively* charged, which causes the leaf to collapse.

However, it is also found that the leaf will collapse when an uncharged object is placed near the cap, and for this reason it is only by observing the

fig 22.7



further divergence of the leaf that the sign of the charge can be definitely established.

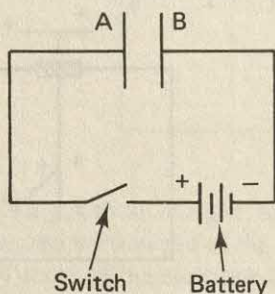
Thus the test for charge is:

If the leaf of a charged electroscope diverges further when an object is placed near its cap, the object has a charge of the same sign as that given to the electroscope

22.6 CHARGING BY USING A BATTERY OR POWER SUPPLY

Conductors may also be charged by using a battery or power supply. Fig. 22.8 shows a battery connected to a switch and two metal plates A and B. When the switch is closed, electrons are removed from plate A making it

fig 22.8 charging by using a battery



positively charged, and are deposited on plate B making it negatively charged. As will be discussed in more detail in Section 23.1, this momentary flow of electrons constitutes an electric current and the flow of electrons can be recorded using an ammeter.

The amount of charge that can be transferred in this way depends on the voltage of the battery, the larger the voltage the larger the amount of charge which is transferred.

It is important to stress at this stage that:

All methods of charging, whether by friction, induction or by use of a battery, involve the movement of electrons. Loss of electrons causes objects to become positively charged, whereas objects that gain electrons become negatively charged

22.7 IONISATION

Air and any other gas is a good insulator as it contains virtually no free electrons. However, it is possible for a gas to become ionised.

Ionisation involves the removal or the gain of an electron by an atom or molecule. If the atom or molecule loses an electron it then has more protons than electrons and it is therefore positively charged. An atom which has lost one or more of its electrons is called a *positive ion*. An atom which has gained electrons is a *negative ion*.

When a gas has been ionised, the electrons are free to move and the gas will conduct. The conduction is different from that which occurs in a metal in that as well as the electrons being free to move, the positive and negative ions will also move. Thus:

Conduction in a gas involves the movement of positive and negative ions and electrons, whereas in a metal it only involves the movement of electrons

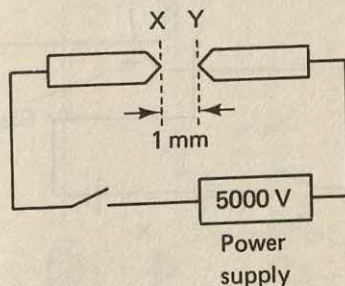
22.8 DISTRIBUTION OF CHARGE ON A CONDUCTOR

When conductors become charged, the charge spreads out over the surface of the conductor. In the case of a sphere, the charge spreads evenly over its outer surface, but for irregularly shaped objects it is found that the charge tends to concentrate in particular places.

This concentration of charge is most pronounced around any sharp point, and causes very strong electrical forces to exist near the point. This can result in the surrounding air becoming ionised.

Fig. 22.9 shows two such pointed objects connected by a switch to a

fig 22.9 *when the switch is closed, a spark occurs between X and Y as the air becomes ionised*



power supply, and when the switch is closed the objects become oppositely charged. The air between them becomes ionised and a spark passes between them provided the voltage is large enough and the objects are placed sufficiently close together.

22.9 LIGHTNING AND LIGHTNING CONDUCTORS

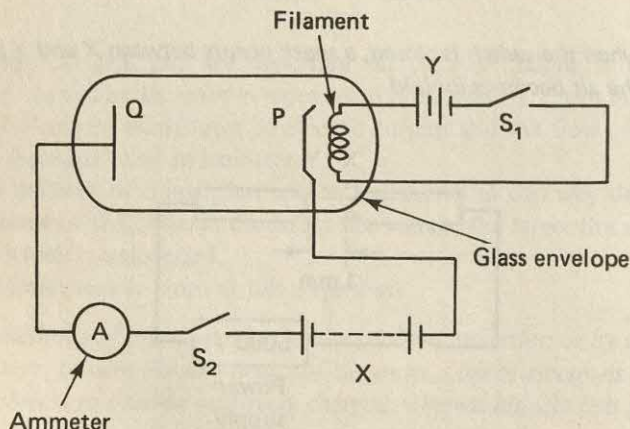
Lightning is caused by the build-up of electric charge on the base of a cloud and a lightning flash results when charge flows between the cloud and the Earth. If this very large flow of charge occurs through a building or a tree, it can cause serious damage, and for this reason buildings are usually protected by using a lightning conductor.

This consists of a pointed spike attached to a thick piece of copper which itself is connected to a large metal plate in the ground. The charge on the base of the cloud induces charge on the Earth beneath it. A high concentration of charge occurs at the top of the conductor and the surrounding air becomes ionised. Because of this ionisation the charge on the base of the cloud is reduced and a lightning flash becomes less likely. Even if a flash does occur, it is more likely to occur through the conductor than through the building and so the building is protected from damage.

22.10 EVIDENCE FOR THE EXISTENCE OF ELECTRONS

Fig. 22.10 shows a glass envelope which is evacuated. At one end there is a small coil of wire, called the filament, and very close to this there is a metal plate called the cathode (P). At the other end of the tube there is another plate called the anode (Q). P and Q are connected to a battery X, a switch

fig 22.10



S_2 and an ammeter which will record any flow of charge that occurs between P and Q. The filament is connected to a separate battery Y and a switch S_1 . When switch S_2 is closed, no current is recorded on the ammeter. However, when S_1 is also closed, a current is recorded, but the current

again falls to zero if the terminals of the battery X are reversed. Therefore, to obtain a current both S_1 and S_2 must be closed and the metal plate Q must be connected to the positive terminal of the battery.

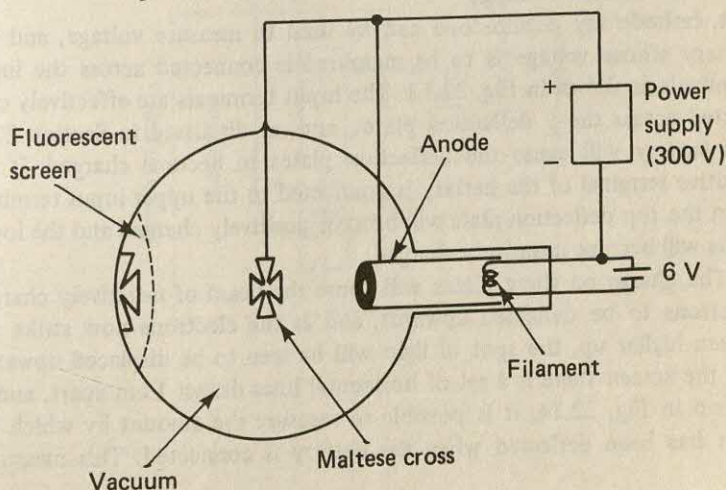
These experimental results can be accounted for as follows. With the switch S_1 closed, the filament gets very hot and heats up the metal plate P. This causes some of the free electrons in the plate to gain sufficient energy to escape from the metal (a process called *thermionic emission*), and these electrons are attracted to the plate Q. This causes a flow of electrons from P to Q and a current is recorded on the ammeter. When Q is connected to the negative terminal of the battery the electrons, whilst still being given off from the plate P, are not attracted to Q, and no current is recorded.

Although originally it was considered that some form of radiation (called *cathode rays*) was being emitted from the metal, it is now known that this 'radiation' consists solely of negatively charged electrons, and this experiment gives evidence that metals contain electrons which, if given sufficient energy, can escape from the metal.

22.11 THE MALTESE CROSS EXPERIMENT

This experiment is illustrated in Fig. 22.11. At the right-hand end of the tube is a filament connected to a 6 volt supply. Close to this there is a cylindrical anode with a small hole in the end further from the filament. A piece of metal in the form of a Maltese cross is fixed opposite this hole. A piece of metal in the form of a Maltese cross is fixed opposite this hole.

fig 22.11 the Maltese cross experiment shows that electrons travel in a straight line in a vacuum



Both the cross and the anode are connected to the positive terminal of a power supply. The end of the tube is coated with a fluorescent material which glows when electrons strike it. With the filament alone switched on, the shadow of the Maltese cross resulting from the light emitted from the filament can be seen on the screen.

When the power supply is switched on the electrons emitted from the filament are attracted to the anode and some of the electrons pass through the hole. These electrons strike the fluorescent screen and produce an 'electron shadow' identical to that produced by the light. The fact that the electrons produce a shadow the same as that produced by the light shows that the electrons travel in straight lines down the tube.

When a magnet is placed near the tube, the shadow produced by the electrons becomes distorted whereas the optical shadow does not change. This indicates that a beam of electrons is deflected in a magnetic field. It is also possible to deflect a beam of electrons by using an electric field and this will be discussed in Section 22.12(b).

22.12 THE CATHODE RAY OSCILLOSCOPE

(a) Introduction

This instrument makes use of a fine beam of fast-moving electrons which travel down an evacuated tube, through two sets of deflection plates, and strike a fluorescent screen. As shown in Fig. 22.12 this fast-moving beam of electrons is produced at one end of the evacuated tube by an 'electron gun', and where the electrons strike the screen a small spot of light is produced. The main controls of the instrument are shown in Fig. 22.13.

(b) Measurement of voltage

The cathode ray oscilloscope can be used to measure voltage, and the battery whose voltage is to be measured is connected across the input terminals as shown in Fig. 22.13. The input terminals are effectively connected across the y deflection plates, and, as discussed in Section 22.6, the battery will cause the deflection plates to become charged. If the positive terminal of the battery is connected to the upper input terminal, then the top deflection plate will become positively charged and the lower plate will become negatively charged.

The charge on these plates will cause the beam of negatively charged electrons to be deflected upwards, and as the electrons now strike the screen higher up, the spot of light will be seen to be displaced upwards. On the screen there is a set of horizontal lines drawn 1 cm apart, and as shown in Fig. 22.14, it is possible to measure the amount by which the spot has been deflected when the battery is connected. This measured

fig 22.12 diagram showing the essential features of a cathode ray tube

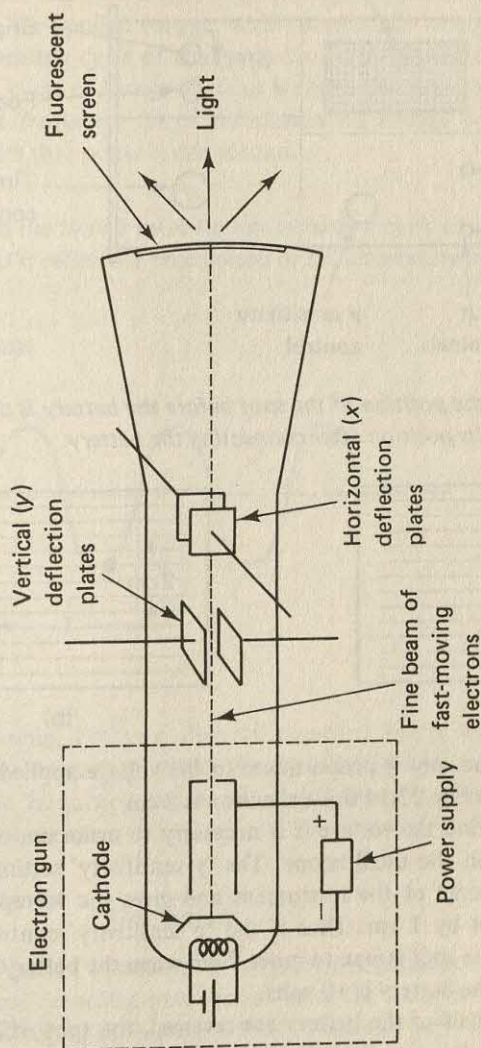


fig 22.13 diagram showing the usual position of the main controls on the front panel of a cathode ray oscilloscope

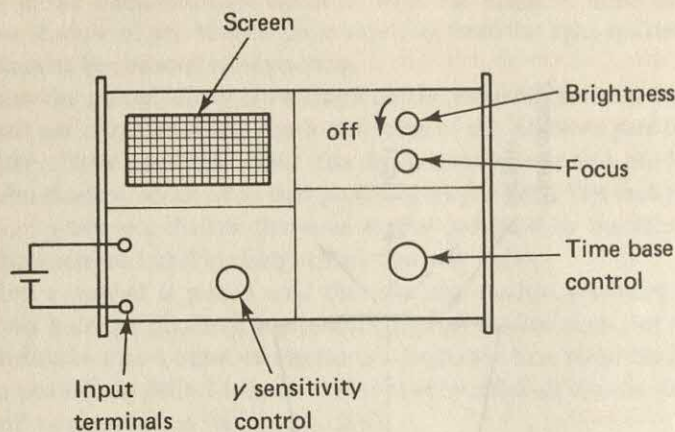
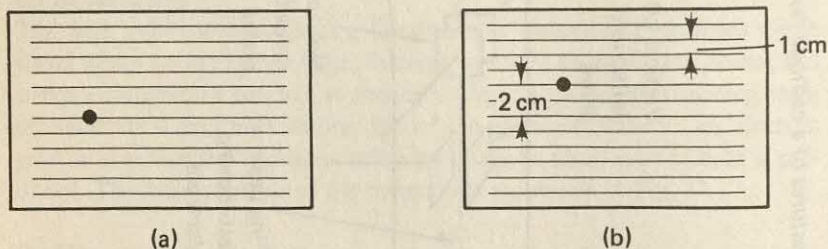


fig 22.14 (a) the position of the spot before the battery is connected and (b) its position after connecting the battery



deflection of the spot is proportional to the voltage applied to the input terminals and in Fig. 22.14 this deflection is 2 cm.

In order to find the voltage it is necessary to make use of the 'y sensitivity' setting on the oscilloscope. The 'y sensitivity' setting can be read off the front panel of the instrument and gives the voltage required to deflect the spot by 1 cm. Thus if the 'y sensitivity' control is set at 5 volts/cm, and the spot is seen to move 2 cm when the battery is connected, the voltage of the battery is 10 volts.

If the terminals of the battery are reversed, the spot will be deflected downwards by the same amount, as now the lower deflection plate will become positively charged and the upper plate negatively charged.

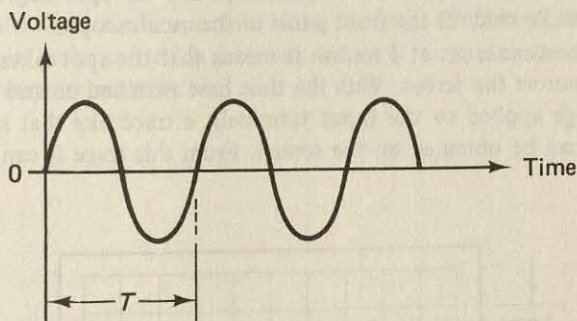
(c) Alternating voltage

A battery gives a steady voltage and the spot will remain fixed in position on the screen once the battery has been connected to the input terminals.

If, however, an alternating voltage is applied to the input terminals, it is found that the spot moves continually up and down on the screen, and this indicates that the voltage applied to the input terminals is continually changing.

The fact that the spot is seen to go through a repeating cycle of change indicates that the applied voltage, whilst continually changing, does itself perform a repeating cycle of change as shown in Fig. 22.15. The time for one complete cycle of change to occur is called the *time period* or *periodic time* (T). The *frequency* (f) of the alternating voltage is the number of complete cycles that occur in one second.

fig 22.15 T is the time it takes for one complete cycle of change to occur and is called the time period of the alternating voltage



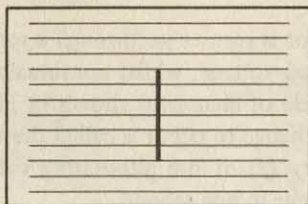
If, for example, $T = 0.1$ s, then 10 complete cycles will occur in one second and the frequency of the supply is 10 cycles/s or 10 hertz (Hz). In general, the frequency can be calculated from the time period using the following equation:

$$\text{Frequency} = \frac{1}{\text{Time period}} \quad (22.1)$$

If the time period is sufficiently large, it is possible to measure it using a stopwatch and recording how long it takes for the spot on the screen to move from the top of the screen to the bottom and back to the top. From this value, the frequency of the alternating voltage can be calculated using equation (22.1).

However, if the time period is very small, the spot will be moving up and down so quickly that a straight vertical line will appear on the screen, as shown in Fig. 22.16. In order to measure the frequency in this case, it is necessary to use the time base control.

fig 22.16 *the trace obtained on an oscilloscope when an alternating voltage of high frequency is applied to the input terminals*

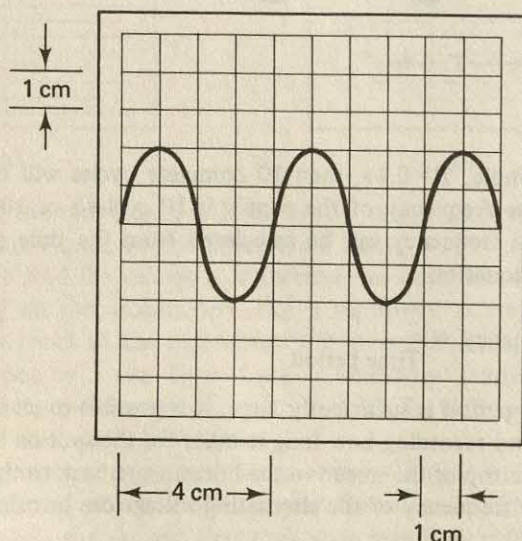


(d) The time base control

The time base control causes the spot to move horizontally across the screen. This is achieved by applying a gradually increasing voltage to the horizontal x deflection plates. The rate at which the spot moves across the screen can be read off the front panel of the oscilloscope.

If the time base is set at 1 ms/cm it means that the spot takes 1 ms to move 1 cm across the screen. With the time base switched on and an alternating voltage applied to the input terminals, a trace like that shown in Fig. 22.17 can be obtained on the screen. From this trace it can be seen

fig 22.17



that the voltage undergoes one complete cycle of change in the time it takes the spot to move 4 cm across the screen. If the time base is set at 1 ms/cm, then the voltage undergoes one complete cycle in 4 ms. Using

equation (22.1) the frequency of the alternating voltage can be found to be 250 Hz.

Thus, as well as measuring voltage, a cathode ray oscilloscope can be used to measure the frequency of an alternating voltage. It may also be used to study wave forms. Fig 22.17 shows one common wave form, but others may be seen by connecting different signals to the input terminals when the time base is switched on.

22.13 TELEVISION

The screen on a television set is the fluorescent screen of a cathode ray tube. The more electrons that strike any one spot the brighter the picture at that point. A special circuit makes the electron beam 'scan' the screen, that is, makes the beam move across the screen in a series of lines from left to right, each line being a little below the one above (the motion is rather like that of the eyes in reading the page of a book). In this way a picture is built up which basically consists of many small patches of light of varying intensity.

ELECTRIC CIRCUITS

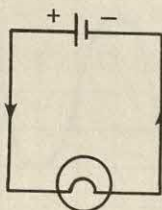
23.1 ELECTRIC CURRENT

Conductors contain a large number of electrons which are free to move within them. When a conductor is connected to a battery, these free electrons, instead of moving in a random way, move in a particular direction.

An electric current is a flow of electrons around an electric circuit

Fig. 23.1 shows a simple electric circuit consisting of a battery, a lamp and two connecting wires. In this circuit, electrons flow from the negative to the positive terminal of the battery. As each electron has a negative

fig 23.1

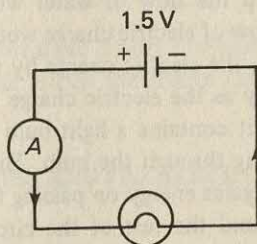


charge, an electric current consists of a flow of negative charge from the negative to the positive terminal of the battery.

However, originally the flow of charge was considered to be from the positive to the negative terminal of the battery. This convention is still internationally followed when representing the direction of the current in a circuit, as is indicated by the arrows in Fig. 23.1, which show the flow occurring from the positive to the negative terminal of the battery. This convention is adopted throughout this book.

Electric current is measured using an ammeter. In order for the ammeter to measure the current, the charge must flow through it, and this is achieved by placing the ammeter in the circuit as shown in Fig. 23.2. Current is measured in *amperes* (A), and the ammeter is said to be connected in series. A reading of 0.2 A on the ammeter would mean that 0.2 coulombs (C) of electric charge pass through the meter in 1 s.

fig 23.2



The value of the current recorded on an ammeter equals the number of coulombs of electric charge that pass through the meter in one second.

$$1 \text{ A} = 1 \text{ C/s} \quad (23.1)$$

If 20 C of charge were to flow through an ammeter in 5 s what current would it record? (Answer = 4 C/s.)

The flow of charge round an electric circuit is similar to the flow of water round a central heating system, and just as the flow of water would be measured in gallons/second so the flow rate of electric charge is measured in coulombs/second.

23.2 ELECTRIC POTENTIAL DIFFERENCE (VOLTAGE)

The battery in Fig. 23.2 is labelled 1.5 volts (V). What does this tell us about the battery?

Before answering this question it is necessary to look more closely at the circuit of Fig. 23.2. The arrows show the flow of positive charge from the positive to the negative terminal of the battery. This positive charge also flows through the battery from its negative to its positive terminal. In order for this to happen, the positive charge must be supplied with energy in order to overcome the electrical forces acting on it.

The voltage of a battery gives the number of joules of energy that are supplied to each coulomb of charge as it passes through the battery

Thus, a 1.5 V battery supplies 1.5 J of energy to each coulomb of charge that passes through it

$$1 \text{ volt} = 1 \text{ joule/coulomb} \quad \text{or} \quad 1 \text{ V} = 1 \text{ J/C} \quad (23.2)$$

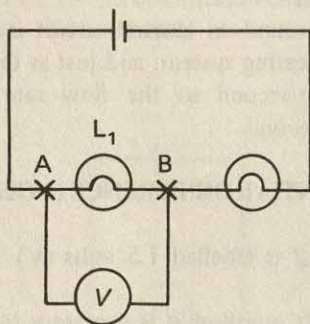
A battery converts chemical energy to electric potential energy. A battery is 'flat' when it has used up its store of chemicals and can no longer supply energy to move the charge through the battery. At this point electric charge ceases to flow round the circuit, and the current drops to zero.

The pump in a water circuit is equivalent to the battery in an electric circuit. Without the pump the flow of water would stop, and similarly without the battery the flow of electric charge would stop.

The energy supplied to the electric charge by the battery is converted into other forms of energy as the electric charge flows round the circuit. For example, if the circuit contains a light bulb the energy is converted to heat and light on passing through the bulb. Thus as the electric charge flows round the circuit, it gains energy on passing through the battery, and gives it up on passing round the rest of the circuit, only to repeat the process over and over again.

Fig. 23.3 shows a voltmeter connected across one of two lamps, lit by a battery. If the voltmeter records 0.6 V it means that each coulomb of electric charge will *lose* 0.6 J of energy as heat and light on passing through the bulb, L_1 . Thus at point A in the circuit the coulomb has 0.6 joules

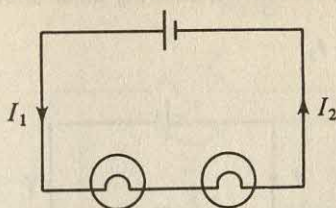
fig. 23.3



more electric potential energy than it does at point B. The voltmeter thus records the difference in the electric potential energy of 1 coulomb between points A and B, and for this reason the reading of the voltmeter is referred to as the potential difference (voltage) between A and B.

23.3 SERIES AND PARALLEL CIRCUITS

Fig. 23.4 shows two bulbs connected in series to a battery and Fig. 23.5 shows the bulbs connected in parallel to the battery. As is clear from the

fig 23.4 a series circuit $I_1 = I_2$ 

diagrams, in a series circuit there is only one path by which the charge can flow, whereas in a parallel circuit the charge can follow *one* of a number of different paths.

23.4 CURRENT IN SERIES AND PARALLEL CIRCUITS

(a) Series

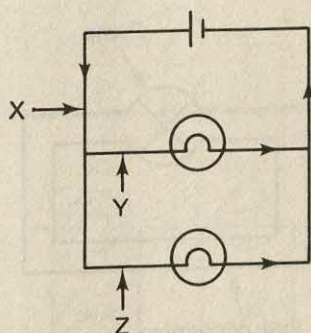
The current is the same at all points in a series circuit

If this were not the case then charge would either be lost from, or build up in, the circuit, neither of which is observed to occur.

(b) Parallel

When the current is measured at points X, Y and Z in the parallel circuit shown in Fig. 23.5, it is found that the current at X equals the sum of the currents at Y and Z. Remembering that current is a flow of charge, this

fig 23.5 a parallel circuit



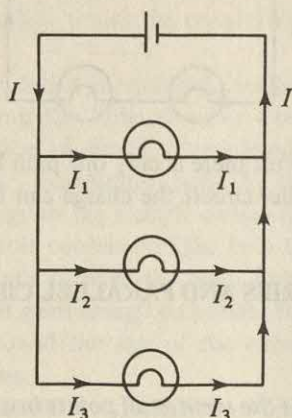
result is to be expected. If the flow rate at X is 6 C/s and at Y is 4 C/s, then the remaining 2 C/s that pass X must flow through Z, so the current at Z is 2 C/s or 2 A. In general, for the parallel circuit shown in Fig. 23.6

$$I = I_1 + I_2 + I_3$$

(23.3)

The sum of the currents flowing in the branches of a parallel circuit equals the current entering those branches

fig 23.6 $I = I_1 + I_2 + I_3$

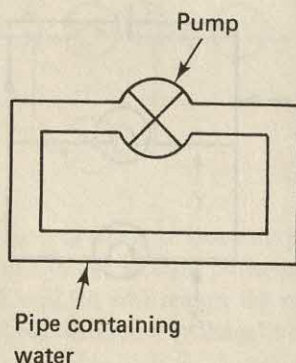


(c) Similarity with flow of water

Both of these results can best be understood by remembering that the flow of charge is similar to the flow of water.

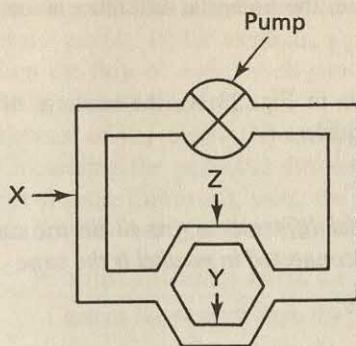
Fig. 23.7 shows a pump connected to a pipe containing water. When the pump is switched on, the water must flow at the same rate at all points in the pipe. This is equivalent to the series circuit.

fig 23.7



Similarly Fig. 23.8 shows the equivalent parallel circuit, and if there is a flow rate of 6 gallons/second at X and 4 gallons/second at Y, then the remaining 2 gallons/second must flow past Z.

fig 23.8



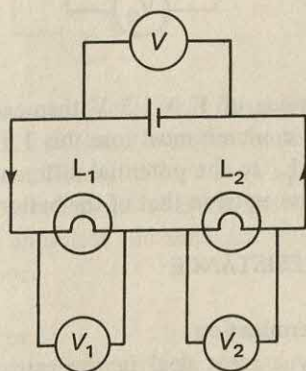
23.5 POTENTIAL DIFFERENCE IN SERIES AND PARALLEL CIRCUITS

(a) Series

For the series circuit shown in Fig. 23.9 it is found that the reading on V is always equal to the sum of the readings on V_1 and V_2 :

$$V = V_1 + V_2 \quad (23.4)$$

Thus the sum of the potential difference across the various components of a series circuit equals the potential difference applied to that circuit

fig 23.9 $V = V_1 + V_2$ 

This result can be explained as follows. If the voltage of the battery in Fig. 23.9 is 6 V then 6 J of energy are supplied to each coulomb that passes through it. If the reading on V_1 is 1 V, then each coulomb loses 1 J

of energy on passing through L_1 , and so must lose the remaining 5 J on passing through L_2 , i.e. the potential difference across L_2 must equal 5 V.

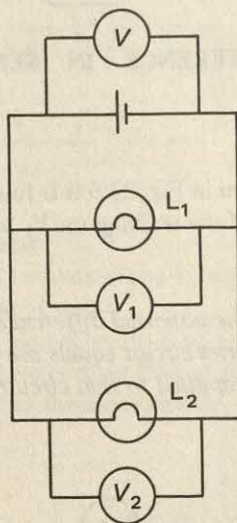
(b) Parallel

For the circuit shown in Fig. 23.10, the readings of V , V_1 and V_2 are always found to be equal:

$$V = V_1 = V_2 \quad (23.5)$$

The potential difference across all electric components connected in parallel is the same

fig 23.10 $V = V_1 = V_2$



In Fig. 23.10 if the value of V is 1.5 V then each coulomb gains 1.5 J from the battery. The coulomb must lose this 1.5 J of energy whether it passes through L_1 or L_2 so the potential difference across both must be the same and in this case equal to that of the battery.

23.6 ELECTRICAL RESISTANCE

(a) Definition and determination

Different objects vary a great deal in the ease with which they allow electrons to flow through them. If a battery is connected across the 'lead' of a pencil, a smaller current will flow than when the same battery is connected across a piece of copper. The pencil 'lead' is said to have a greater electrical resistance than the copper.

It is important to realise that the pencil 'lead' will limit the flow of charge at *all points* in a circuit. This is equivalent to placing an obstruction of some sort in a water circuit. If, for example, a pad of cotton wool is secured in a pipe, then the flow of water at *all points* in the pipe will be lower than if the cotton wool were removed.

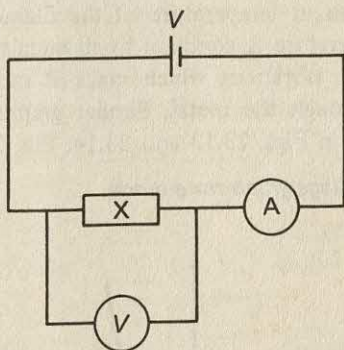
The electrical resistance of any object (X) can be found by connecting it to a battery and measuring the potential difference across it, and the corresponding current flowing through it, using the circuit shown in Fig. 23.11. The resistance can then be calculated from the defining equation:

$$\text{Resistance (ohms)} = \frac{\text{Potential difference across the object (volts)}}{\text{Current flowing through the object (amps)}}$$

or

$$R = \frac{V}{I} \quad (23.6)$$

fig 23.11 a circuit for measuring the resistance of X

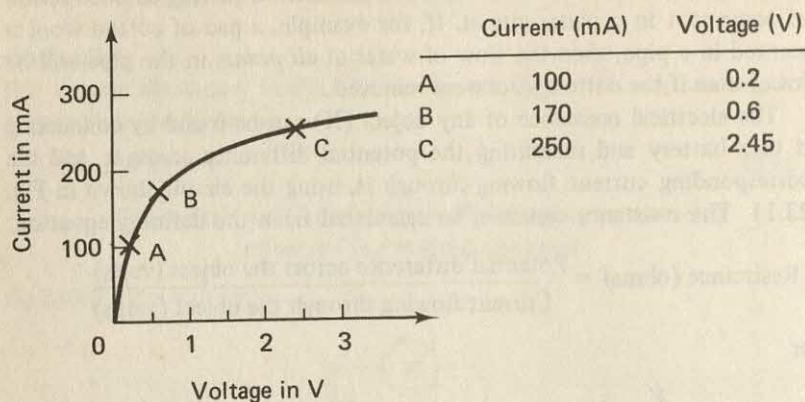


If the potential difference is measured in volts and the current in amps, then the resistance will be calculated in ohms. An object has a resistance of 1 ohm (Ω) if a potential difference of 1 volt (V) gives rise to a current of 1 amp (A). The electrical resistance measures the voltage needed to pass a current of 1 A through an object and the higher the resistance the greater this voltage will have to be.

(b) Current/voltage curves

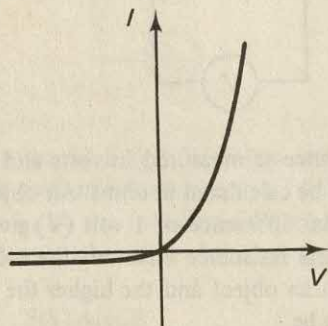
In general, the electrical resistance of an object is not constant and will vary with the potential difference that is applied to it. An ordinary filament lamp does not have a constant resistance and this variation can be seen from a current/voltage graph for a typical lamp (Fig. 23.12). From the graph the resistance of the lamp at the points A, B and C can be calculated,

fig 23.12 current/voltage graph for a bulb



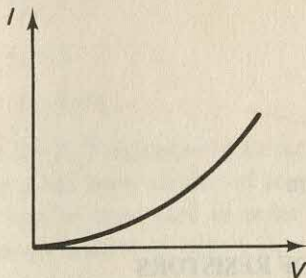
and is found to have the values of $2\ \Omega$, $3.5\ \Omega$, and $9.8\ \Omega$ respectively. The resistance of a filament lamp increases as the potential difference applied to it increases. The increase in the resistance of the filament in the lamp is due to the increase in temperature of the filament. The increase in resistance with temperature is common to all metals and arises owing to the increased atomic vibrations which make it more difficult for the electrons to pass through the metal. Similar graphs for a diode and a thermistor are shown in Figs. 23.13 and 23.14. Fig. 23.13 shows that the

fig 23.13 current/voltage graph for a diode



diode conducts an electric current easily in one direction whilst only allowing a very small current to pass through it in the opposite direction.

The current/voltage graph for the thermistor (Fig. 23.14) curves in the opposite direction to that for a lamp and the electrical resistance of the thermistor decreases as the voltage applied to it increases. Thermistors are made from materials called *semiconductors*, and the resistance of a semi-

fig 23.14 *current/voltage graph for a thermistor*

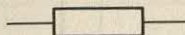
conductor decreases as its temperature increases. The reason for this is that as the temperature of the semiconductor increases, the number of electrons which are free to move in the semiconductor increases and this allows a greater current to pass through it.

(c) A special case — Ohm's law

Certain objects have current/voltage graphs which pass through the origin and are linear over a large range of voltages. For such objects the current is directly proportional to the potential difference applied to them, and they are said to obey Ohm's law, which may be stated as follows:

At a constant temperature the current in a conductor is directly proportional to the potential difference across it

A resistor is an object which obeys Ohm's law, and is manufactured to have some specific value of resistance. In circuit diagrams it is represented by the symbol



If a resistor has a resistance of $4\ \Omega$ it means that a potential difference of $4\ \text{V}$ will give a current of $1\ \text{A}$.

(d) Resistivity

The resistance R of any object depends on its length l , its cross-sectional area A , and on the material of which it is made. The *resistivity* is defined by the following equation:

$$R = \frac{\rho \times l}{A} \quad (23.7)$$

where ρ is a constant which takes account of the material of which the object is made, and is called its resistivity. At a given temperature the resistivity of any material is fixed, but different pieces of the material will have different values for resistance depending on their length and

Rearranging equation (23.7) enables the units for resistivity to be found:

$$\rho = \frac{R \times A}{l}$$

so the units are

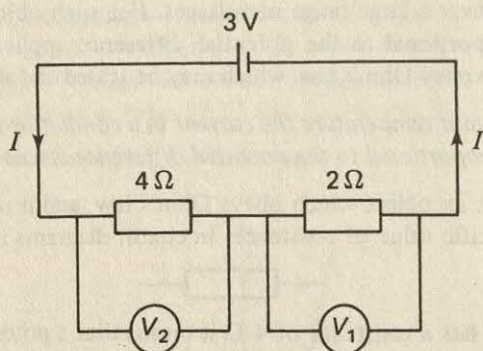
$$\frac{\Omega \times \text{m}^2}{\text{m}} \equiv \Omega \text{ m}$$

23.7 COMBINATION OF RESISTORS

(a) Resistors in series

Fig. 23.15 shows two resistors of values 4Ω and 2Ω connected in series to a 3 V battery. What is the size of the current I in the circuit?

fig 23.15



As the resistors are connected in series, the current in both resistors must be the same, and using equation (23.6) gives

$$2 = \frac{V_1}{I} \quad \text{and} \quad 4 = \frac{V_2}{I}$$

Rearranging these equations gives:

$$V_1 = 2 \times I \quad \text{and} \quad V_2 = 4 \times I$$

As the sum of the potential differences across the components of a series circuit equals the applied potential difference, it follows that:

$$V_1 + V_2 = 3 \text{ V}$$

Thus

$$(2 \times I) + (4 \times I) = 3$$

$$\therefore I(2 + 4) = 3$$

$$I = \frac{3}{6} = \frac{1}{2}$$

The size of the current is $\frac{1}{2}$ A. This is exactly the same size current as would occur if the 3 V battery had been connected across a single 6Ω resistor. Thus a 4Ω and 2Ω resistor connected in series have exactly the same effect in a circuit as a single resistor of value 6Ω .

In general, resistors of value R_1, R_2, R_3 connected in series have the same effect in a circuit as one resistor, whose resistance R is the sum of R_1, R_2 and R_3

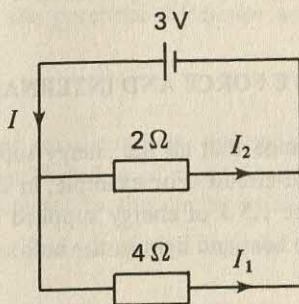
$$R = R_1 + R_2 + R_3 \quad (23.8)$$

(b) Resistors in parallel

Fig. 23.16 shows a 4Ω and a 2Ω resistor connected in parallel to a 3 V battery. As the resistors are connected in parallel, the potential difference across each must be 3 V and using equation (23.6) it follows that:

$$4 = \frac{3}{I_1} \quad \text{and} \quad 2 = \frac{3}{I_2}$$

fig 23.16



Rearranging gives:

$$I_1 = \frac{3}{4} \quad \text{and} \quad I_2 = \frac{3}{2}$$

As the resistors are in parallel

$$I = I_1 + I_2$$

Therefore

$$I = \frac{3}{4} + \frac{3}{2} = 3 \left(\frac{1}{4} + \frac{1}{2} \right)$$

If R is the value of the single resistor that has the same effect in a circuit as the $4\ \Omega$ and $2\ \Omega$ resistors in parallel, it would draw the same current I from the 3 V battery. 3 V across $R\ \Omega$ results in a current of $3/R$ amps.

Therefore

$$\frac{3}{R} = 3 \left(\frac{1}{4} + \frac{1}{2} \right)$$

which gives

$$\frac{1}{R} = \frac{1}{4} + \frac{1}{2} \quad \text{i.e.} \quad R = \frac{4}{3}\ \Omega$$

Thus a single resistor of value $\frac{4}{3}\ \Omega$ has the same effect in a circuit as $4\ \Omega$ and $2\ \Omega$ resistors connected in parallel.

In general if resistors of values R_1 , R_2 and R_3 are connected in parallel, they have the same effect in a circuit as one resistor whose value R is given by the equation:

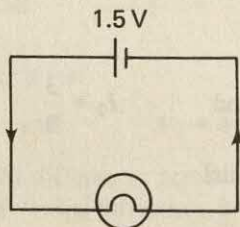
$$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} \quad (23.9)$$

As an example of the use of this equation, show that resistors of values 2 , 4 and $6\ \Omega$ connected in parallel behave as a single resistor of value $1.09\ \Omega$.

23.8 ELECTRO-MOTIVE FORCE AND INTERNAL RESISTANCE

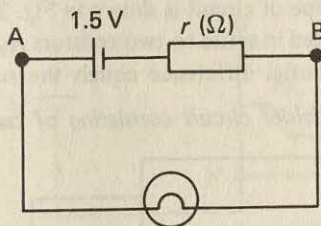
So far it has been assumed that all the energy supplied by the battery is delivered to the external circuit. For example, in the circuit of Fig. 23.17 it is assumed that *all* the 1.5 J of energy supplied by the battery to each coulomb is converted to heat and light in the bulb.

fig 23.17



This, however, is not strictly correct. The battery has some resistance, and some of the energy will be lost as heat in the battery itself as the charge flows through it. The resistance of the battery is called its *internal resistance*. Thus not all the energy supplied by the battery is delivered to the external circuit, and the circuit of Fig. 23.17 should strictly speaking be drawn as shown in Fig. 23.18, where A and B represent the terminals of the battery and r its internal resistance.

fig 23.18



The e.m.f. or electro-motive force, E volts, of a battery is the total energy supplied to each coulomb of charge that passes through it, including any energy that may be lost as heat in the battery itself

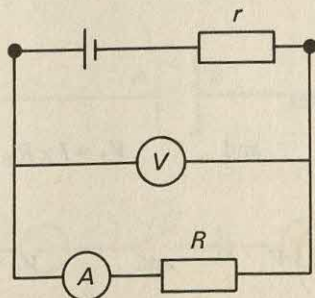
The potential difference, V volts, across the terminals of a battery is the energy which is delivered to the external circuit by each coulomb of charge

The difference between the e.m.f. of the battery and the terminal potential difference is the potential difference across the internal resistance. Hence

$$E - V = Ir \quad (23.10)$$

The value of r can be calculated from this equation provided the values of E , I and V are known. The values of V and I can be found by connecting a voltmeter and an ammeter to the circuit as shown in Fig. 23.19. The

fig 23.19 a circuit for measuring the internal resistance of a battery



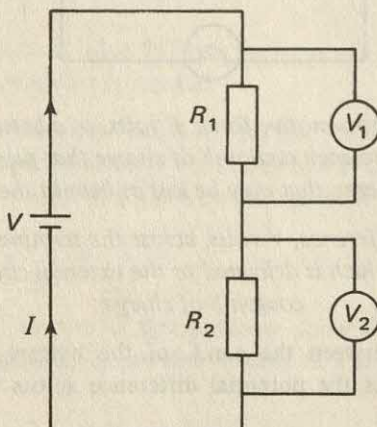
value of E equals the value of V when $I = 0$, and so can be found by disconnecting the resistor (R) from the battery ($I = 0$), and recording the value registered on the voltmeter.

23.9 THE POTENTIAL DIVIDER CIRCUIT

This is a particularly useful circuit as it enables a battery of fixed potential difference, V volts, to be adapted to supply a potential difference of any value from 0 to V volts.

An example of this type of circuit is shown in Fig. 23.20. It simply consists of a battery connected in series to two resistors R_1 and R_2 . For such a circuit the applied potential difference equals the sum of the potential

fig 23.20 a potential divider circuit consisting of two resistors in series



differences across the two resistors. Thus in such a circuit the potential difference of the battery is split or divided between the two resistors. The potential difference is divided between the two resistors in direct proportion to the value of their resistances, the larger value resistor having the larger proportion of the applied potential difference.

In Fig. 23.20 the value of the current I in the circuit is given by:

$$I = \frac{V}{R_1 + R_2}$$

Using equation (23.6) gives,

$$V_1 = I \times R_1 \quad \text{and} \quad V_2 = I \times R_2$$

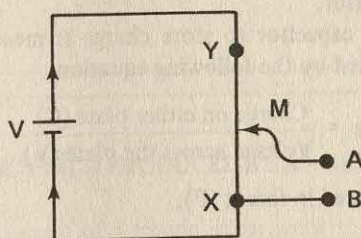
Thus

$$V_1 = \left(\frac{R_1}{R_1 + R_2} \right) V \quad \text{and} \quad V_2 = \left(\frac{R_2}{R_1 + R_2} \right) V$$

Thus by suitably choosing the values of R_1 and R_2 it is possible to obtain any fraction of the applied voltage V across R_2 .

The circuit is more useful if R_1 and R_2 are replaced by a single piece of high resistance wire XY as shown in Fig. 23.21. M is a moveable contact and, as it is moved from point X to point Y on the wire, the value of the

fig 23.21 a potential divider circuit consisting of a high-resistance wire XY



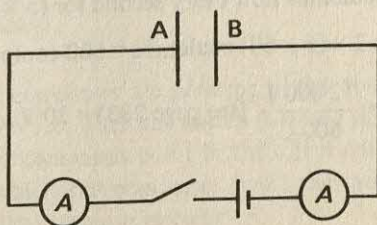
potential difference across AB takes on all values from 0 to V volts. Thus any value of potential difference from 0 to V volts can be obtained across AB , by suitably positioning the contact M on the wire.

23.10 CAPACITORS

A simple capacitor consists of two flat pieces of metal separated by an insulator of very high resistance. One way of achieving this is to use two metal plates with an air gap between them. The circuit shown in Fig. 23.22 consists of such a capacitor connected in series with a switch, two ammeters and a battery.

When the switch is closed, it is found that both ammeters register a current, and that after a short time the current falls to zero. The fact that the ammeters register a current means that there has been a flow of charge through them, and plate B has received negative charge, and plate A has had negative charge (electrons) removed from it, making it positively

fig 23.22



charged. As both meters behave in an identical way, the charge lost from plate A is exactly the same as that gained by plate B. The capacitor is now said to be charged and has a potential difference across it equal to that of the battery and, as a result of this, no further current is recorded on the meters (see Section 22.6).

If the experiment is repeated using a battery of higher voltage, more charge will be transferred from one plate to the other and it is found that the amount of charge transferred is directly proportional to the voltage applied to the capacitor.

The ability of a capacitor to store charge is measured by its capacitance, which is defined by the following equation:

$$\text{Capacitance} = \frac{\text{Charge on either plate (C)}}{\text{Voltage across the plate (V)}}$$

Capacitance is measured in farads (F).

The capacitance measures the amount of charge stored on each plate of the capacitor when a voltage of 1 V is applied across it

A capacitor of 6×10^{-6} F has 6×10^{-6} C of charge transferred from one plate to the other when a voltage of 1 V is applied to it. As the charge transferred and the voltage are directly proportional to each other, it is possible to scale these values to find other corresponding sets of values for charge transferred and voltage, e.g. 10 V would cause 6×10^{-5} C to be transferred from one plate to the other.

In the case of a capacitor consisting of two metal plates, it is found that the capacitance depends on the distance apart of the plates, and also on their area of overlap. A variable capacitor consists of two sets of plates, one of which is fixed, whilst the other can be rotated, and this varies the capacitance by varying the degree of overlap between the plates.

WORKED EXAMPLE

A current of 2 A flows through a resistor for 5 minutes. The heat evolved is 12 kJ. What is the potential difference of the supply?

Current = 2 A $\therefore$ 2 coulombs flow every second for (5×60) s

Charge which flows = $2 \times (5 \times 60)$ coulombs = 600 coulombs

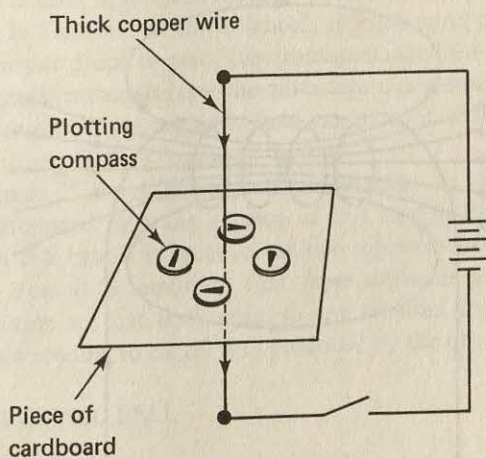
Potential difference = $\frac{12\,000 \text{ J}}{600 \text{ C}}$ (see page 233) = 20 V

MAGNETIC EFFECTS OF AN ELECTRIC CURRENT

24.1 THE MAGNETIC FIELD PRODUCED BY AN ELECTRIC CURRENT

Fig. 24.1 shows a diagram of a simple experiment to demonstrate that a magnetic field is produced by an electric current. It consists of a long piece of thick copper wire connected in series with a switch and a battery. The wire passes through a piece of cardboard on which is placed a number of plotting compasses.

fig 24.1 *an electric current produces a magnetic field*



Initially all the compasses are pointing in the direction of the Earth's magnetic field. However, when the switch is closed this produces a current in the wire and the compasses point in the directions shown in Fig. 24.1. The direction in which the compasses point indicates that the magnetic field produced by the current is circular.

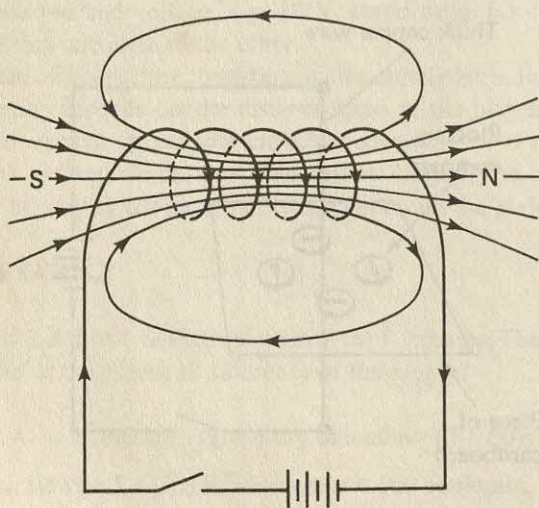
If the terminals of the battery are reversed it produces a current in the opposite direction. In this case, although the same circular pattern is obtained, the compasses point in the opposite direction to that shown in Fig. 24.1. This shows that the direction of the magnetic field is reversed. The direction of the magnetic field produced by the current in a wire can be found using the *right-hand grip rule*.

The wire is gripped with the right hand so that the thumb points in the direction of the current. The direction in which the fingers curl round the wire indicates the direction of the magnetic field

24.2 THE MAGNETIC FIELD PRODUCED BY A SOLENOID

As shown in Fig. 24.2 a solenoid consists of a long piece of wire wound in the shape of a spring or helix. When the switch in Fig. 24.2 is closed, the current in the solenoid produces a magnetic field. If the magnetic field pattern is investigated using iron fillings or plotting compasses, it is found to be similar to that produced by a bar magnet. The ends of the solenoid corresponding to the North and South poles of the bar magnet are labelled N and S in Fig. 24.2.

fig 24.2 the magnetic field produced by a solenoid

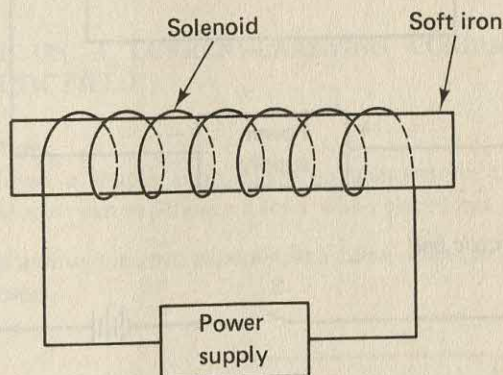


The direction of the field along the axis of the solenoid can be found using the right-hand grip rule, and its strength depends on the size of the current and the number of turns per unit length of the solenoid.

24.3 THE ELECTROMAGNET

An electromagnet consists of a solenoid wound on a piece of soft iron (Fig. 24.3). The solenoid is insulated from the soft iron. When the solenoid is connected to a power supply, the current in it produces a magnetic field. This field magnetises the piece of soft iron and as a result a much stronger field is produced than would occur with the solenoid alone.

fig 24.3 *an electromagnet*



Soft iron is used in an electromagnet because, although it is magnetised when there is a current in the solenoid, it loses most of its magnetism when the current drops to zero. Electromagnets used for lifting and transporting magnetic materials have the advantage over permanent magnets in that the strength of the magnetic field can be controlled by altering the size of the current in the solenoid.

The strength of the magnetic field produced by an electromagnet can be further increased by using a piece of soft iron in the form shown in Fig. 24.4. In this case it is usual to use two solenoids wound on each side of the soft iron. It is important that these solenoids are wound in the correct direction so that the current in one solenoid does not produce a magnetic field tending to cancel that produced by the other solenoid.

24.4 THE ELECTRIC BELL

Fig. 24.5 shows the essential features of an electric bell. It consists of an electromagnet, similar to the type shown in Fig. 24.4, which is positioned near to a small iron bar A. This bar is attached to a springy piece of metal B.

When the switch S is closed, the current follows the path shown by the arrows. This current magnetises the soft iron in the electromagnet and as

fig 24.4

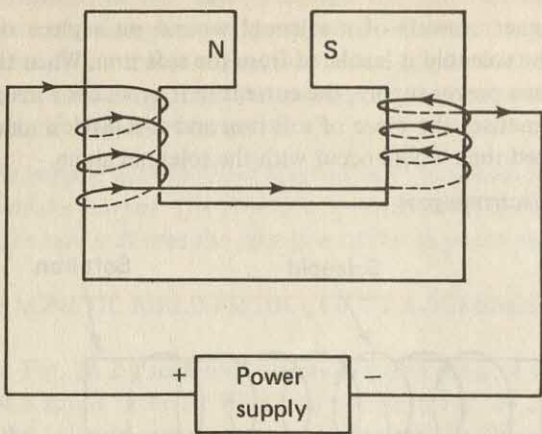
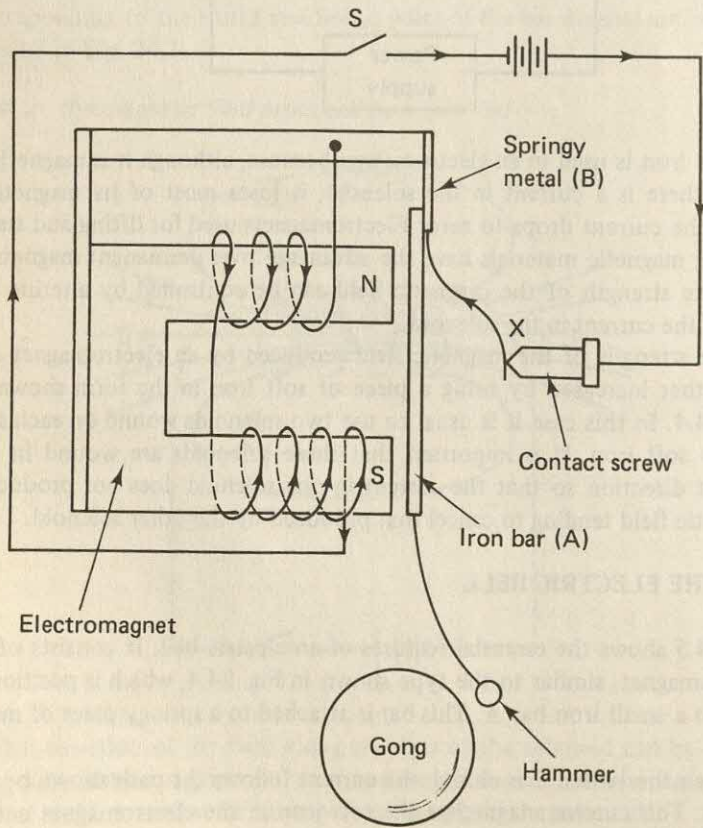


fig 24.5 an electric bell



as a result the iron bar A moves towards the poles of the magnet. This movement causes the hammer to strike the gong.

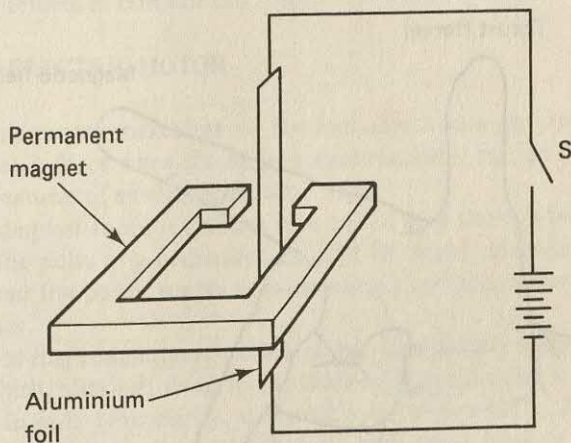
The movement also causes the electric circuit to be broken at the point X in Fig. 24.5, and the current in the circuit drops to zero. As a result the soft iron in the electromagnet ceases to be magnetic and the iron bar A is pulled back to its original position by the springy piece of metal B. This re-establishes contact at the point X and so the whole cycle of change is now repeated. The bell will ring continuously provided the switch S is closed.

24.5 FORCE ON A CURRENT-CARRYING CONDUCTOR IN A MAGNETIC FIELD

(a) Introduction

Fig. 24.6 shows a simple arrangement to demonstrate that a current-carrying conductor can experience a force when placed in a magnetic field.

fig 24.6 *the aluminium strip experiences a force when the switch S is closed*



It consists of a strip of aluminium foil supported between the poles of a permanent magnet and connected in series with a switch and a battery. With the switch open, the aluminium experiences no force as it is not magnetic. However, when the switch is closed a current is produced in the foil, and the foil is seen to move. This indicates that a force is acting on it.

This force arises because of the interaction that occurs between the electric current and the magnetic field and if either the magnet is removed, or the switch is opened, the force drops to zero.

(b) The size of the force

In general it is found that the size of the force produced in this way is proportional to both the size of the current and the strength of the magnetic field. More surprisingly, it is also found that the size of the force depends on the relative *directions* of the electric current and the magnetic field.

The force has its largest value when the magnetic field is applied at right-angles to the direction of the current, but is zero when the direction of the field is the same or opposite to that of the electric current.

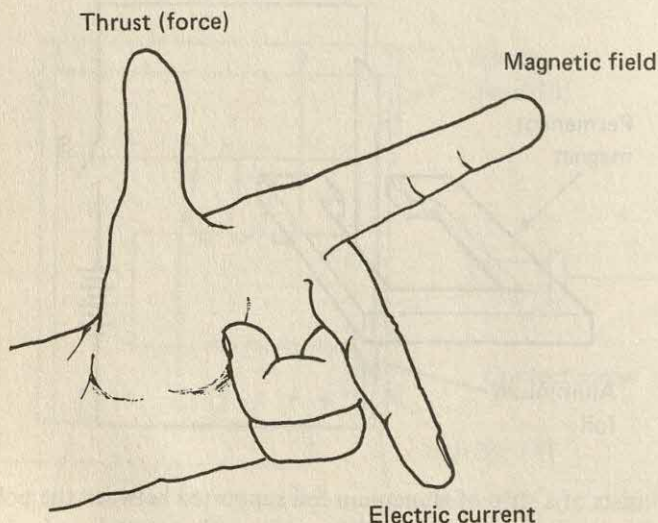
(c) The direction of the force

As well as in part controlling the size of the force, the relative directions of the field and current also determine its direction.

In Fig. 24.6 if either the direction of the current or the direction of the magnetic field is reversed, then the foil will be observed to move in the opposite direction, indicating that the force acting on it has been reversed.

The direction of the force resulting from the interaction of an electric current and a magnetic field can be found using the *left-hand rule* (Fig. 24.7).

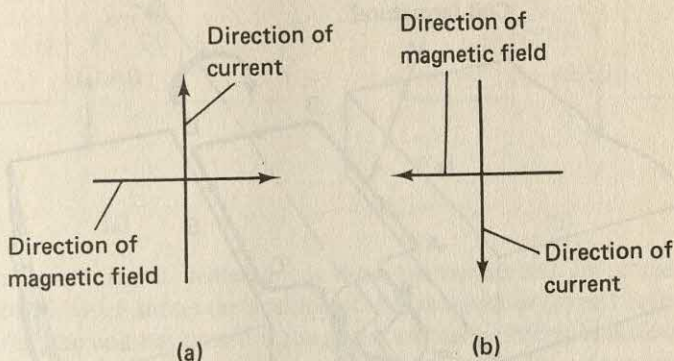
fig 24.7 the left-hand rule



If the thumb, fore-finger and second finger of the left hand are held mutually at right-angles to each other, with the seCond finger pointing in the direction of the Current and the Fore-finger in the direction of the Field, then the THumb points in the direction of the THRust acting on the conductor

Use this rule to show that the direction of the force on the conductor in Fig. 24.8a and b is, in both cases, directed into the plane of the paper.

fig 24.8



(d) The ampere

If a current flows in the same direction in two parallel wires, there is a force between them. The greater the current, the greater is the force. The ampere is defined in terms of this force.

24.6 THE ELECTRIC MOTOR

The electric motor makes use of the fact that a wire carrying a current experiences a force when placed in a magnetic field. Fig. 24.9 shows the essential features of an electric motor.

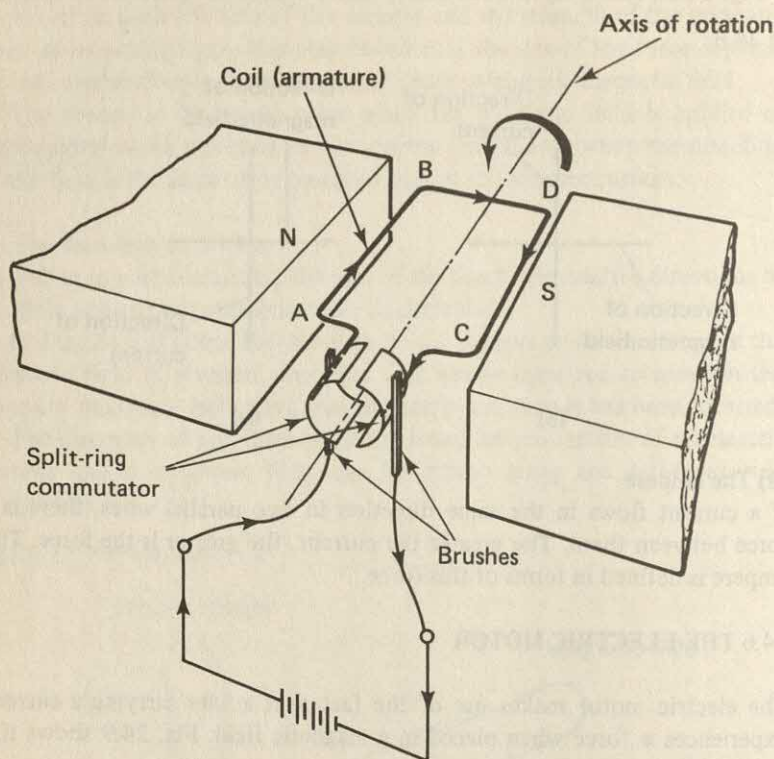
In its simplest form, it consists of a coil of wire which is free to rotate between the poles of a permanent magnet. Electrical connection between the coil and the power supply is made using a *split-ring commutator* and two *brushes*.

The split-ring commutator consists of two cylindrically shaped pieces of copper which, although fixed to the shaft on which the coil is wound, are insulated from it. (For clarity, this shaft is not shown in Fig. 24.9.) These two pieces of copper are insulated from each other because of the gap between them and each strip is connected to one end of the coil.

The brushes can be made from two pieces of wire, each of which is connected to one terminal of the power supply. The brushes are positioned so that they press sufficiently firmly on the split-ring commutator to make good electrical contact with it, whilst still allowing the shaft to rotate.

The current follows the path shown by the arrows in Fig. 24.9 and, as is clear from the diagram, the direction of the current in each side of the coil is at right-angles to the magnetic field. Using the left-hand rule, Section 24.5(c), the side AB of the coil will experience a downwards

fig 24.9 an electric motor



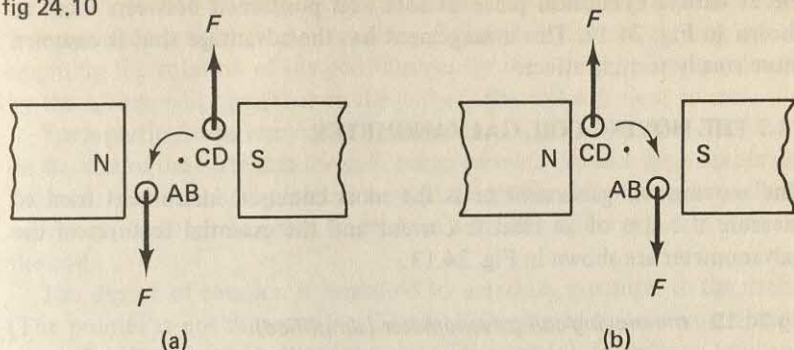
force, and the side CD an upwards force. As a result of these forces, the coil will rotate in the direction shown in Fig. 24.9.

However, as the coil passes through the vertical position, it is necessary that the direction of the force on each side of the coil is reversed. The reason for this can be understood by considering Fig. 24.10.

Fig. 24.10a shows the forces F acting on the coil just before it passes through the vertical position and Fig. 24.10b shows the situation just after it has passed through the vertical position. As is clear from these diagrams, if the forces on each side of the coil continue to act in the same direction, the turning effect they produce will change direction and cause the coil to stop rotating. Thus, in order to keep the coil rotating in one direction, it is necessary to reverse the direction of the forces acting on it each time it passes through the vertical position.

As was discussed in Section 24.5(c), the direction of the force is reversed if the current changes direction. The split-ring commutator produces exactly this effect as it causes the direction of the current in the coil to reverse each time it passes through the vertical position.

fig 24.10



The way this is achieved can best be understood by considering Fig. 24.11 which shows the position of the commutator (a) just before and (b) after the coil has passed through the vertical position. As is clear from these diagrams, the electrical contact to each side of the commutator P and Q is reversed each time the coil passes through the vertical position. This causes the direction of the current in the coil to reverse and as a result of this the coil continues to rotate in the same direction.

In commercial motors a number of coils rotate in a *radial magnetic field*. Such a field is produced by using a magnet with cylindrically shaped pole

fig 24.11 the commutator reverses the direction of the current in the coil

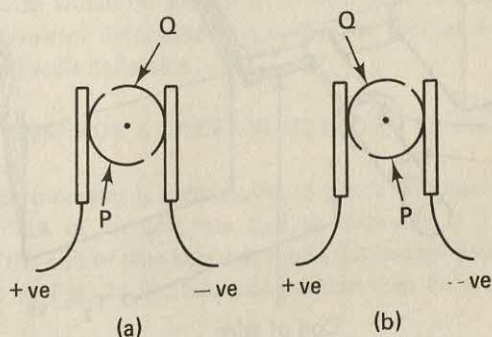
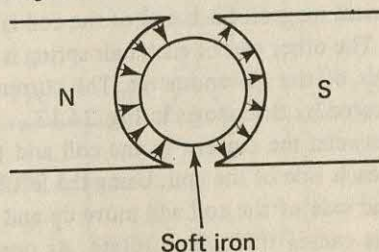


fig 24.12 a radial magnetic field

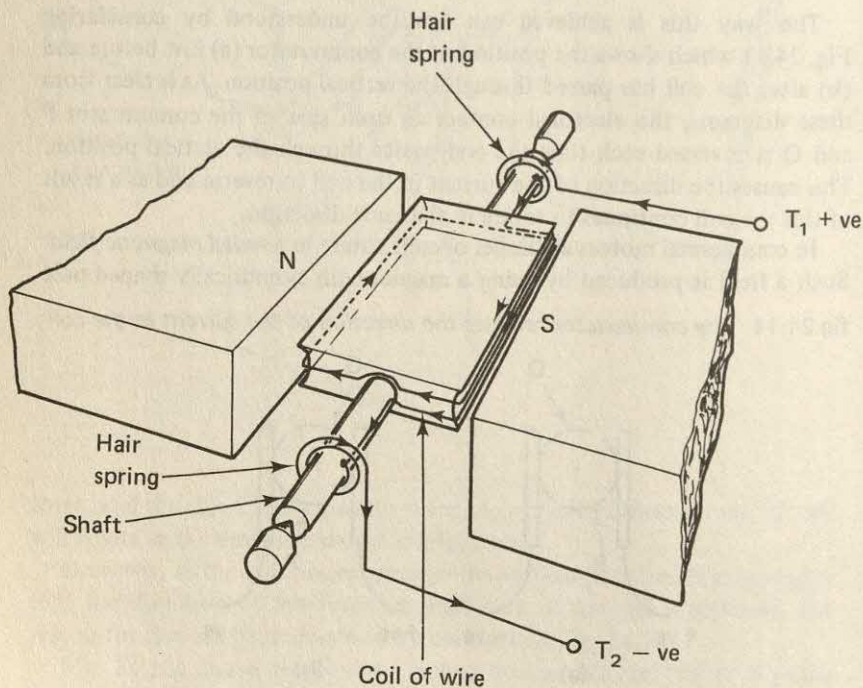


pieces with a cylindrical piece of soft iron positioned between them as shown in Fig. 24.12. This arrangement has the advantage that it causes a more steady turning effect.

24.7 THE MOVING-COIL GALVANOMETER

The moving-coil galvanometer is the most common instrument used to measure the size of an electric current and the essential features of the galvanometer are shown in Fig. 24.13.

fig 24.13 *the moving-coil galvanometer (simplified)*



It consists of a coil of wire wound on a shaft which is pivoted between the poles of a permanent magnet. Each end of the coil is connected to one end of a hair spring. The other end of each hair spring is connected to one of the input terminals of the galvanometer. The current to be measured follows the path indicated by the arrows in Fig. 24.13.

The interaction between the current in the coil and the magnetic field produces a force on each side of the coil. Using the left-hand rule, Section 24.5(c), the right-hand side of the coil will move up and the left-hand side will move down. This causes the coil to rotate. As one end of each hair

spring is attached to the shaft, the rotation of the coil and shaft causes the hair springs to tighten. As a result of this they exert a turning effect opposing the rotation of the coil. Eventually the turning effect produced by the springs will equal that on the coil and the coil will come to rest.

For a particular galvanometer the amount of rotation produced depends on the size of the current in the coil. Large currents produce large forces on the coil, Section 24.5(b), and as a result a large amount of rotation will be necessary for the hair spring to produce a turning effect equal to that on the coil.

The degree of rotation is measured by attaching a pointer to the shaft. (The pointer is not shown in Fig. 24.13.) This pointer moves across a scale which is calibrated to indicate the size of the current. In order to produce a linear scale (that is, one with equal spacing between the divisions), a radial magnetic field must be used (Fig. 24.12).

When the galvanometer is disconnected from the circuit, the current in the coil drops to zero. Thus the turning effect produced by the hair springs will return the coil to its original position and the pointer will return to zero.

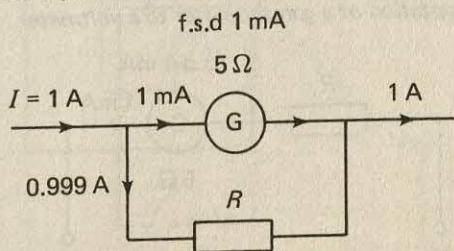
The size of the current in the coil which produces the maximum or *full-scale deflection* (sometimes abbreviated to f.s.d.) depends on the construction of the galvanometer. If a full-scale deflection is to be caused by a small current, then a strong magnet, weak hair springs and a coil having a large number of turns would be used in its construction. Thus the construction of the galvanometer determines the size of the current in the coil which produces a full-scale deflection.

24.8 CONVERSION OF A GALVANOMETER TO AN AMMETER

Suppose a galvanometer is constructed to give a full-scale deflection for a current of 1 mA in the coil, and that its resistance is $5\ \Omega$ (that is, the resistance of the coil of wire through which the current passes is $5\ \Omega$).

As shown in Fig. 24.14, this galvanometer can be adapted to record

fig 24.14 *the adaptation of a galvanometer to record larger values of current by putting a shunt across the galvanometer*



higher values of current by placing a low resistance shunt in parallel with it. The size of the resistor R determines the size of the current I that is required to produce a full-scale deflection.

Suppose this current I is to be 1 A. In order to produce a full-scale deflection, the current in the coil of the galvanometer *must be* 1 mA. Using equation (23.3) the current in the resistor R must be 0.999 A. The resistor is chosen to allow this precise value of current to pass through it. The size of this resistor can be calculated in the following way:

Using equation (23.6) the potential difference V_g across the galvanometer is given by the equation:

$$V_g = (1 \times 10^{-3} \times 5) \text{ volts}$$

and the potential difference V_r across the resistor is given by:

$$V_r = (0.999 \times R) \text{ volts}$$

As the resistor and the galvanometer are in parallel and using equation (23.5), it follows that:

$$V_g = V_r$$

$$\therefore (1 \times 10^{-3} \times 5) = (0.999 \times R)$$

This gives the value of R to be $5 \times 10^{-3} \Omega$.

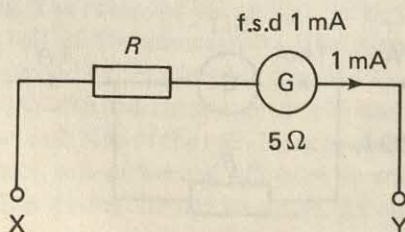
Thus, by placing a resistor of $5 \times 10^{-3} \Omega$ in parallel with the galvanometer in Fig. 24.14, it is possible to adapt it to record currents up to a maximum value of 1 A.

24.9 CONVERSION OF A GALVANOMETER TO A VOLTMETER

Fig. 24.15 shows the same galvanometer adapted for use as a voltmeter. This is achieved by placing a high resistor R in series with the galvanometer.

The value of the resistor is chosen to limit the current in the galvanometer to 1 mA when the maximum potential difference to be measured is applied across the points marked X and Y in Fig. 24.15. The value of the resistor depends on the size of this potential difference and its value can be calculated in the following way:

fig 24.15 the adaptation of a galvanometer to a voltmeter



Suppose this maximum voltage is to be 1 V, then using equation (23.6) the total resistance between X and Y has to be $10^3 \Omega$ in order to limit the current in the galvanometer to 1 mA. This resistance is made up of the resistance of the galvanometer (5Ω) and the resistor (R). Using equation (23.8) it follows that:

$$10^3 = 5 + R$$

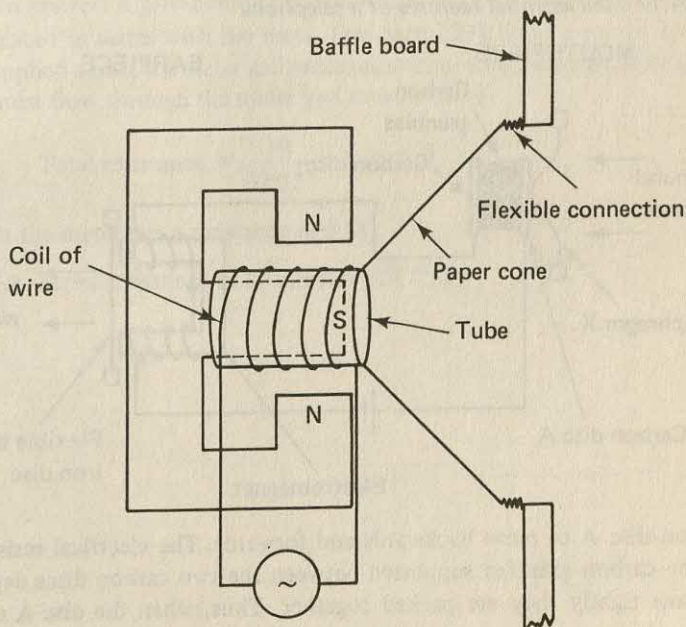
$$\therefore R = 995 \Omega$$

Thus, by placing a resistor of value 995Ω in series with the galvanometer in Fig. 24.15, it is adapted for use as a voltmeter to record potential differences up to a maximum value of 1 V.

24.10 THE MOVING-COIL LOUDSPEAKER

As is shown in Fig. 24.16, a moving-coil loudspeaker consists of a coil of wire wound on a tube. One end of the tube is fixed rigidly to a paper cone, and the other end of this cone is loosely supported in a circular hole in a piece of board. This board is called the baffle board. The flexible connection with the board means that the cone, tube and coil can move backwards and forwards along the axis of the tube. The coil itself is positioned in a radial magnetic field.

fig 24.16 *the moving-coil loudspeaker*



When an alternating voltage is applied to the coil it produces an alternating current. The interaction between the current in the coil and the magnetic field causes the coil to move and when the current changes direction the coil will move in the opposite direction (Section 24.5(c)).

Thus, the alternating current in the coil causes it to move backwards and forwards along its axis and the paper cone performs a similar motion. The movement of the cone produces compressions and rarefactions in the air and a sound wave is thus produced. The frequency of the sound wave is identical to the frequency of the alternating voltage.

If the size of the current in the coil is increased, the force acting on it will be larger (Section 24.5(b)). This larger force will produce a greater movement of the cone and coil and the loudness of the sound wave will increase.

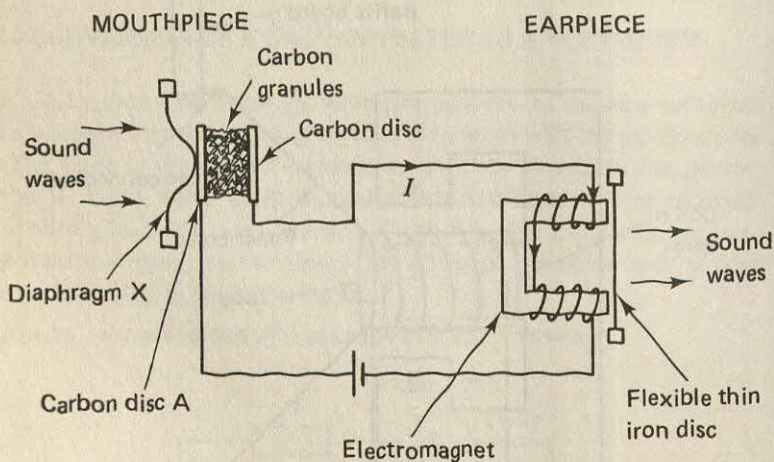
Thus, the moving-coil loudspeaker converts electrical energy from the voltage supply into sound energy.

24.11 THE TELEPHONE

The earpiece of a telephone behaves in a similar way to the moving-coil loudspeaker in that a change in an electric current in a coil causes sound waves to be produced.

The essential features of the mouthpiece and earpiece of a telephone are shown in Fig. 24.17. Sound waves striking the diaphragm X cause the

fig 24.17 *the essential features of a telephone*



carbon disc A to move backwards and forwards. The electrical resistance of the carbon granules supported between the two carbon discs depends on how tightly they are packed together. Thus, when the disc A moves

backwards and forwards the electrical resistance across the granules alters and the current I changes. This change in the current alters the strength of the magnetic field produced by the electromagnet in the earpiece (see Section 24.3), and hence the force on the thin iron disc making up the diaphragm in the earpiece alters. As a result of this, the iron disc moves backwards and forwards producing a sound wave.

WORKED EXAMPLE

A moving coil meter has a resistance of $5\ \Omega$. It gives a full-scale deflection when a current of $15\ \text{mA}$ passes through it. How may it be converted into (i) an ammeter reading 0 to $10\ \text{A}$ and (ii) a voltmeter reading 0 to $10\ \text{V}$?

- (i) When $10\ \text{A}$ arrives at the meter, $15\ \text{mA}$ ($0.015\ \text{A}$) must pass through the meter and $9.985\ \text{A}$ must pass through the shunt. If the shunt has a resistance of $R\ \Omega$ the p.d. across the shunt is $9.985 \times R$ volts (see page 239).

$$\text{Potential difference across meter} = 0.015 \times 5 = 0.075\ \text{V}$$

But these two p.d.s. are the same since the resistors are in parallel. Hence

$$9.985 R = 0.075$$

$$R = 0.00751\ \Omega$$

- (ii) To convert a galvanometer into a voltmeter, a high resistance must be placed in series with the meter (see page 239). When a p.d. of $10\ \text{V}$ is applied across the meter and series resistance, then a current of $0.015\ \text{A}$ must flow through the meter and series resistor.

$$\text{Total resistance} = \frac{10}{0.015} = 666.7\ \Omega$$

But the meter has a resistance of $5\ \Omega$

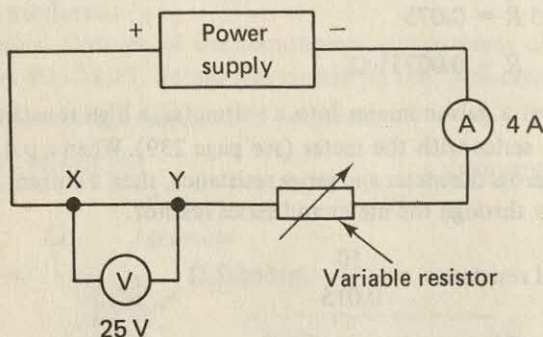
$$\text{Series resistance to be added} = 661.7\ \Omega$$

ENERGY AND POWER FROM ELECTRICAL SOURCES

25.1 ELECTRICAL POWER AND ENERGY

Fig. 25.1 shows a power supply connected in series to a small piece of resistance wire XY, a variable resistor and an ammeter. The potential difference across the wire can be measured using the voltmeter.

fig 25.1 *an electric current produces a heating effect*



When the supply is switched on the wire becomes hot, and for this to occur it must be being supplied with energy. The rate at which energy is supplied to the wire, that is the energy supplied to it in one second, can be calculated from the values of the current and potential difference recorded on the meters.

If the potential difference across the wire is 25 V, then each coulomb of charge loses 25 J of electric potential energy on passing down the wire (see Section 23.2). This energy is converted to heat energy in the wire. If the current is 4 A, then 4 C of charge pass down the wire in 1 s (see Section 23.1), and as *each* coulomb supplies 25 J of energy, a total of 100 J is supplied to the wire *in one second*. Thus, the wire is supplied with energy at the rate of 100 J/s.

The rate at which energy is supplied is defined as the *power*, and for any electrical component the power is calculated using the following equation:

$$\begin{array}{lcl} \text{Power} & = & \text{Current } (I) \times \text{Potential difference } (V) \\ \text{(in watts)} & & \text{(in amps)} \quad \text{(in volts)} \end{array} \quad (25.1)$$

Using equation (23.6) the potential difference across any component is given by the equation:

$$\text{Potential difference } (V) = \text{Current } (I) \times \text{Resistance } (R)$$

and substituting this result into equation (25.1) gives:

$$\text{Power} = I^2 \times R \quad (25.2)$$

If the value of the variable resistor in Fig. 25.1 is decreased, the readings on the ammeter and voltmeter will increase. According to equation (25.1) the power will be greater, and this is confirmed by the fact that the wire becomes hotter, indicating that energy is being supplied to it at a greater rate.

The *total* amount of energy supplied to any electrical component depends on the rate at which it is supplied with energy and on how long it is connected to the supply. Thus, if the wire in Fig. 25.1 is supplied with energy at the rate of 100 J/s and it is connected to the supply for 100 s, a total of 10 000 J will be supplied to the wire. In general the total amount of energy supplied to any electrical component is calculated using the following equation:

$$\begin{array}{lclcl} \text{Energy} & = & \text{Current} & \times & \text{Potential difference} & \times & \text{Time} \\ \text{(J)} & & \text{(A)} & & \text{(V)} & & \text{(s)} \end{array} \quad (25.3)$$

25.2 POWER USING AN ALTERNATING VOLTAGE

When a source of alternating voltage (see Section 22.12(c)) is applied to an electric circuit, it will cause a continually changing or alternating current in the circuit. It will also give rise to a continually changing or alternating potential difference across each component in the circuit. Fig. 25.2 shows a source of alternating voltage connected across a piece of resistance wire XY, and a variable resistor, and Fig. 25.3 shows how the potential difference across the wire and the current in it vary with time.

The maximum values of the potential difference and current are called their peak values

In Fig. 25.3 the peak values of the potential difference and current are 25 V and 4 A respectively.

fig 25.2

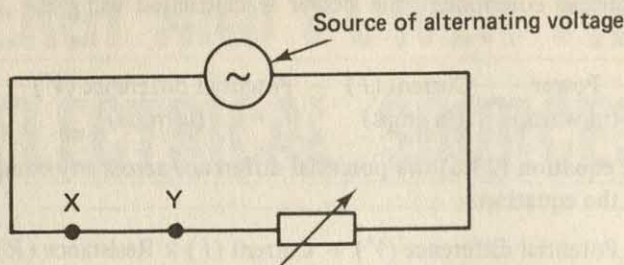
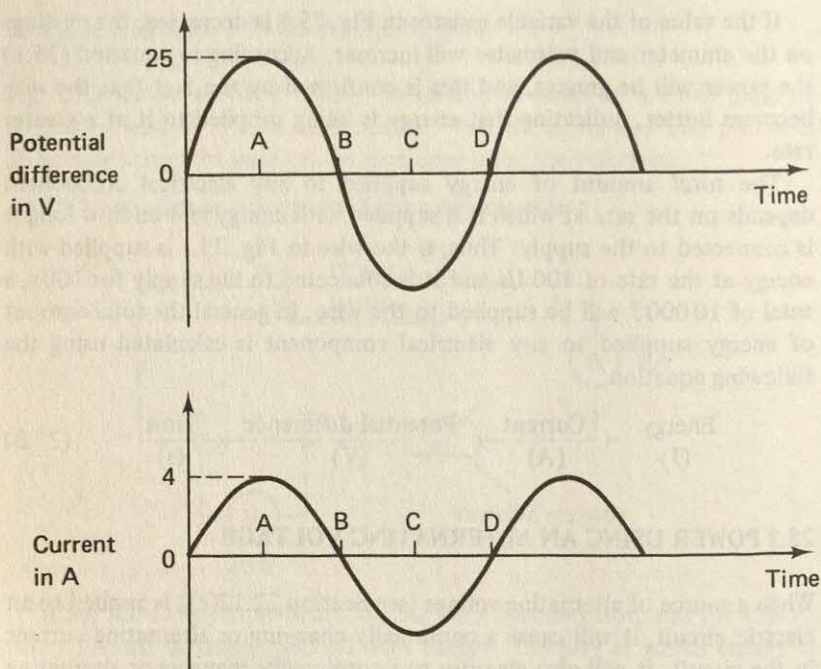


fig 25.3 the peak value of the potential difference and current are 25 V and 4 A respectively



As both the potential difference and current are continually changing it follows from equation (25.1) that the rate at which energy is being supplied to the wire must also be continually changing. At the times marked A and C in Fig. 25.3, energy is being supplied to the wire at its maximum rate of 100 J/s. At other times the power will be less than this, and at the times marked O, B and D, the power will momentarily be zero.

In order to calculate the *average rate* at which energy is supplied to the wire, it is necessary to use the r.m.s. (root mean square) values of the

potential difference (V) and current (I) ($V_{\text{r.m.s.}}$ and $I_{\text{r.m.s.}}$). Equation (25.1) can be rewritten as follows:

$$\begin{array}{c} \text{Power} \\ (\text{J/s or W}) \end{array} = \begin{array}{c} V_{\text{r.m.s.}} \\ (\text{V}) \end{array} \times \begin{array}{c} I_{\text{r.m.s.}} \\ (\text{A}) \end{array} \quad (25.4)$$

Unless stated to the contrary, it can be assumed that the values given for an alternating potential difference or an alternating current refer to the r.m.s. values and *not* the peak values.

25.3 THE MAINS SUPPLY

In most countries the electricity supply to houses uses an alternating voltage. For example, in Britain the alternating voltage is supplied at 240 V and has a frequency (see Section 22.12(c)) of 50 Hz.

All domestic appliances in Britain are designed to be used off this mains supply and are marked for use at 240 V. Also marked on the appliance is its power rating and this enables the current in the appliance to be calculated using equation (25.4). For example, a light bulb marked 240 V, 100 W would draw a current of 0.42 A (from $100 = 240 \times I$) and the current in a 240 V, 2 kW electric fire would be 8.3 A.

It is important to appreciate that the power rating of the appliance has the value stated *only* when connected to a 240 V supply, and if it is connected to a source of voltage less than 240 V it will be supplied with less energy every second.

25.4 DOMESTIC WIRING

As shown in Fig. 25.4 the mains supply is delivered to a house using two wires called the *live* and *neutral*, which usually enter the house by means of an underground cable. These wires then supply separate circuits within the house. A number of these circuits are used for lighting; others, called the *ring mains*, supply the electric points or sockets; and a separate circuit is used to supply the cooker. Fig. 25.4 shows a circuit of each type.

Separate circuits are used because of the different power ratings of the appliances which will be connected to them, and because of this the wiring in each circuit is designed to take a certain maximum value of current. If the current exceeds this value the wires could overheat and cause a fire. In order to avoid this, each circuit has a fuse connected in the live wire which 'blows' when the current exceeds a certain value. A fuse is simply a piece of wire which melts when the current reaches a certain value and thus prevents any further flow of current in the circuit. The fuses connected to each circuit are contained within a single fuse box (not shown in

fig 25.4 domestic wiring

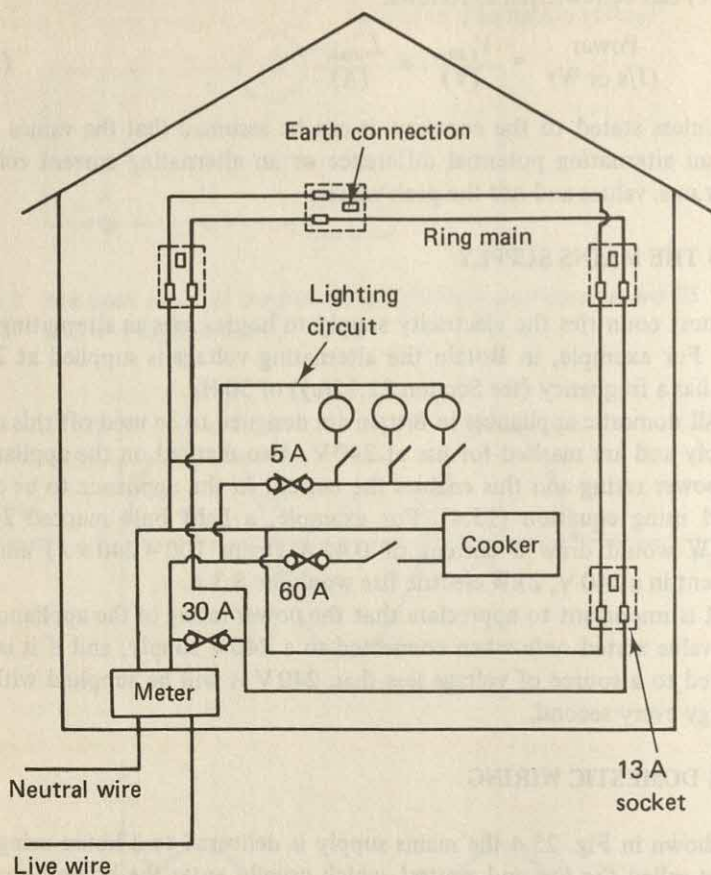


Fig. 25.4), and when replacing a fuse it is most important to use the correct rating of fuse wire.

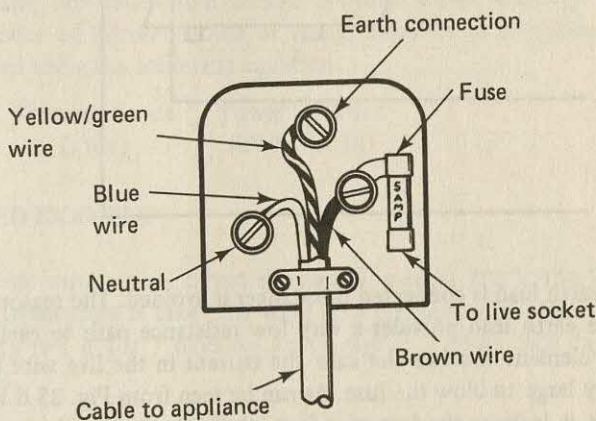
As shown in Fig. 25.4, within each separate circuit appliances are connected across the live and neutral wires and each appliance is controlled by a switch connected in the live wire.

As well as allowing connection to the live and neutral wires all points, apart from those used for lighting, also enable a connection to be made to earth. This is achieved by means of a third wire (not shown in Fig. 25.4), which is usually connected to a water pipe. Water pipes come up through the ground and make good connections to earth. As will be discussed in Section 25.6, this earth lead is necessary purely as a safety measure and does not normally carry a current.

25.5 ELECTRICAL CONNECTION TO THE MAINS SUPPLY

Most domestic appliances are connected to the mains supply by means of a three-pin plug (Fig. 25.5). This enables the appliance to be connected to

fig 25.5 a three-pin plug



the live, neutral and earth wires and it is most important that the wires from the appliance are connected in the correct way using the following colour code:

Brown wire (Live); Blue wire (Neutral); Green/Yellow wire (Earth)

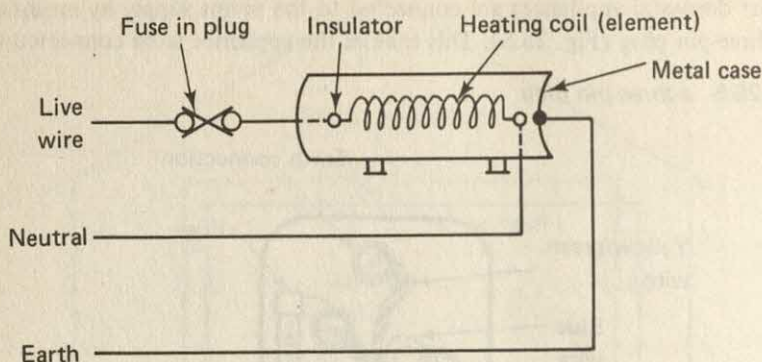
The plug also contains a fuse connected in the live wire. The current rating of this fuse depends on the power rating of the appliance which is being connected to the mains. The fuse prevents damage to the appliance which could occur if excessively large currents passed in it.

25.6 THE EARTH LEAD

As was mentioned in Section 25.4, the earth lead is required purely as a safety precaution. To illustrate this, consider the electric fire shown in Fig. 25.6. One end of the heater coil or element is connected, via the fuse in the plug, to the live wire and the other end is connected to the neutral wire. The surrounding metal case is connected to the earth and is insulated from the element.

However, if the element breaks it is possible that the broken end, connected to the live wire, will touch the metal case. Without the earth connected, anyone touching the case would receive an electric shock as it is in contact with the live wire.

fig 25.6 the earth wire connected to the case is a safety device



If the earth lead is connected this danger is avoided. The reason for this is that the earth lead provides a very low resistance path to earth. Thus, when the element touches the case the current in the live wire becomes sufficiently large to blow the fuse. As can be seen from Fig. 25.6 when the fuse blows it isolates the live wire from the element and the metal case, and prevents any danger of an electric shock.

The reason the lighting circuits do not have an earth lead is that even if the filament breaks the broken end connected to the live wire will only come into contact with the glass envelope of the bulb. As glass does not conduct, the bulb can be touched without any danger of an electric shock.

Therefore, it is only important to connect an earth lead when the outer case of the appliance is made of a conducting material.

25.7 THE KILOWATT-HOUR

As with all power supplies, the mains supply is a source of energy, and as discussed in Section 25.3, the power rating of any domestic appliance gives the rate at which it is supplied with energy when connected to the mains. The total amount of energy supplied can be found from the power rating and the time for which the appliance is connected to the mains.

The meter, connected across the mains supply (Fig. 25.4) records the total amount of energy supplied to the house. The energy consumption is recorded on the meter in kilowatt-hours.

One kilowatt-hour (kWh) is the total amount of energy supplied to a 1 kW appliance when it is connected to the mains supply for 1 h

As a 1 kW appliance is supplied with energy at the rate of 1000 J/s, a total of 3.6×10^6 J of energy will be supplied in 1 h (3600 s). Thus:

$$1 \text{ kWh} = 3.6 \times 10^6 \text{ J of energy}$$

It is important to appreciate that the kilowatt-hour is a unit of energy and that the payment to the Electricity Board is based on the amount of energy used. At present in Britain (1982) one kilowatt-hour of energy costs about 4 pence. A 2 kW fire left on for 4 h will cost the same as a 100 W lamp left on for 80 h, as both consume 8 kWh of energy. In general, the number of kilowatt-hours of energy supplied to an appliance can be calculated using the following equation:

$$\begin{array}{l} \text{Kilowatt-hours} \\ (\text{kWh}) \end{array} = \begin{array}{l} \text{Power} \\ (\text{kW}) \end{array} \times \begin{array}{l} \text{Time} \\ (\text{h}) \end{array} \quad (25.5)$$

WORKED EXAMPLE

An electric motor on a carpet sweeper is rated at 250 V 150 W. (i) How much current does it take? (ii) What is its resistance? (iii) How much does it cost to run it for 7 days for an average of 2 hours each day? (1 kWh costs 6p).

(i) Power = Current $\times$ Potential Difference (page 265)

$$150 = \text{Current} \times 250$$

$$\text{Current} = \frac{150}{250} = 0.6 \text{ A}$$

(ii) $V = IR$ (page 239)

$$250 = 0.6 \times R$$

$$R = \frac{250}{0.6} = 417 \Omega$$

(iii) Number of kWh = $0.15 \times (7 \times 2) = 2.1$

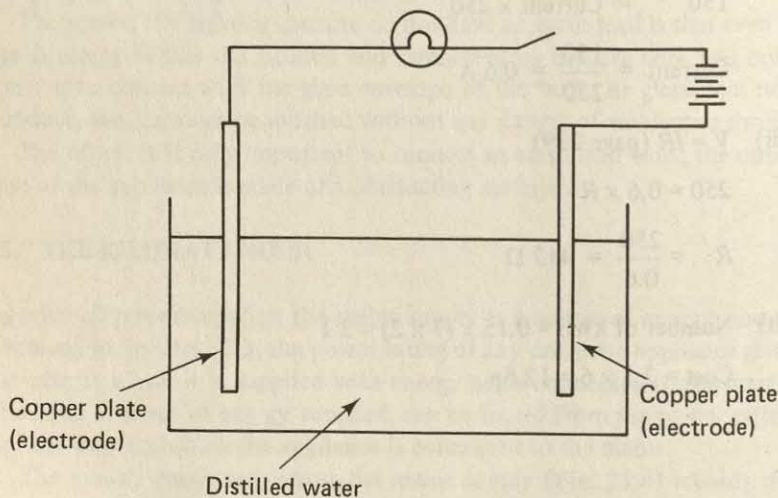
$$\text{Cost} = 2.1 \times 6 = 12.6\text{p}$$

THE CHEMICAL EFFECT OF AN ELECTRIC CURRENT

26.1 INTRODUCTION

Fig. 26.1 shows two copper plates, called *electrodes*, immersed in some distilled water. The plates are connected in series with a bulb, a switch and a battery. When the switch is closed the bulb does not light as distilled water is a poor conductor.

fig 26.1 *distilled water is a good insulator as it contains very few free ions*

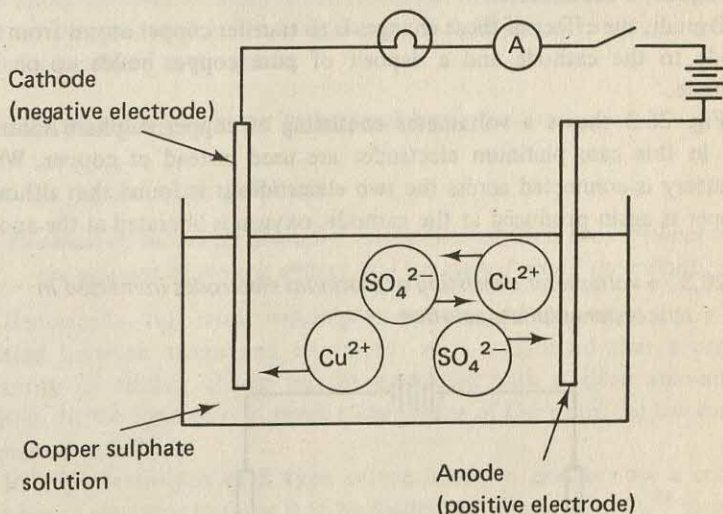


However, if some copper sulphate crystals are dissolved in the water a current flows in the circuit and the bulb lights. The reason for this difference in behaviour between the copper sulphate solution and the distilled water can be explained in the following way.

A solution of copper sulphate contains the ions Cu^{2+} and SO_4^{2-} . The copper ion Cu^{2+} is a copper atom which has lost two electrons and as a

result has an overall positive charge. The two electrons lost by the copper atom are gained by the sulphate ion, SO_4^{2-} , which results in this ion possessing an overall negative charge. These ions are free to move around in a random way within the solution, and this behaviour is similar to the random movement of the free electrons within a metal. However, when a battery is connected across the two electrodes immersed in the solution, the ions, instead of moving in a random way, start to move in particular directions. Fig. 26.2 shows the directions in which the ions move. The copper ions move towards the negative electrode, called the *cathode*, and the sulphate ions travel towards the positive electrode, which is called the *anode*. The flow of the charged ions in the solution causes an electric current in the circuit.

fig 26.2 *electrolysis of copper sulphate solution using copper electrodes*



Liquids such as copper sulphate solution which contain free positive and negative ions are called *electrolytes*. Strong electrolytes contain a large number of free ions, whereas weak electrolytes contain only a few free ions.

Distilled water is a very weak electrolyte. The very few free ions that do exist in the distilled water are produced by the *dissociation* of the water molecules themselves into hydrogen ions, H^+ , and hydroxide ions, OH^- . As the number of these ions is extremely small their effect is negligible and distilled water behaves as an insulator.

26.2 ELECTROLYSIS

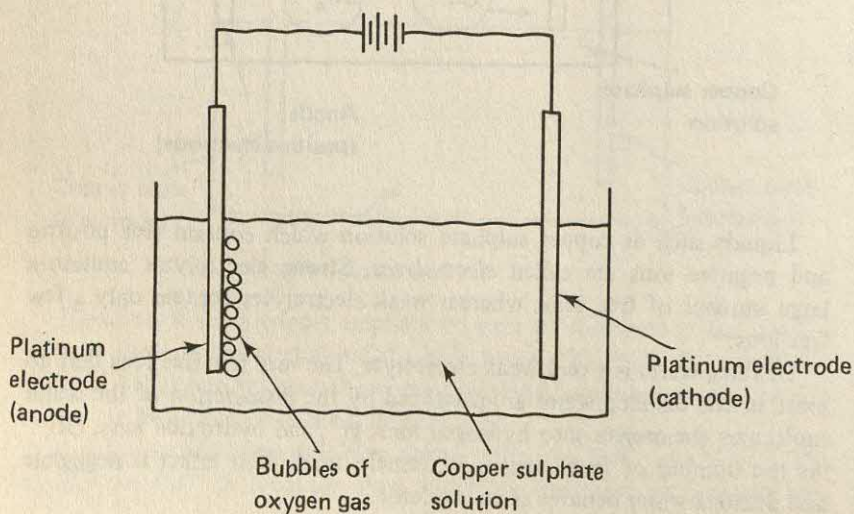
The passage of an electric current through an electrolyte causes chemical changes to occur at each electrode and is called *electrolysis*. A tank containing an electrolyte with two electrodes immersed in it is called a *voltameter*.

As an example of the chemical changes that occur at the electrodes, consider the voltameter shown in Fig. 26.2. The electric current in the circuit causes electrons to flow onto the cathode and to be removed from the anode. The positive Cu^{2+} ions in the electrolyte are attracted to the negative electrode. Electrons which flow onto the cathode combine with the Cu^{2+} ions to produce neutral copper atoms which are deposited on the cathode. At the anode, copper atoms in the electrode are converted into Cu^{2+} ions and two electrons are released. The electrons pass round the circuit to the cathode whilst Cu^{2+} ions move through the solution from the anode to the cathode.

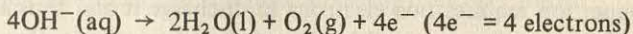
Overall, the effect of these changes is to transfer copper atoms from the anode to the cathode and a deposit of pure copper builds up on the cathode.

Fig. 26.3 shows a voltameter consisting of copper sulphate solution but in this case platinum electrodes are used instead of copper. When a battery is connected across the two electrodes it is found that although copper is again produced at the cathode, oxygen is liberated at the anode.

fig 26.3 *a voltameter consisting of platinum electrodes immersed in copper sulphate solution*



In order to explain the production of oxygen gas at the anode it is necessary to consider the negative ions which are attracted to it. The copper sulphate solution contains the two negative ions SO_4^{2-} and OH^- (the OH^- ions resulting from the dissociation of the water molecules). On arriving at the anode either of these ions could lose electrons, but as the OH^- ions lose their electrons more easily than the SO_4^{2-} ions, the change that occurs at the anode may be summarised as follows:



The four electrons released in this way pass round the circuit to the cathode and oxygen gas bubbles off from the anode.

These two examples of electrolysis show that the chemical changes that occur at the electrodes depend on both the nature of the electrodes and on the nature of the electrolyte. However, in general, the change occurring at the cathode involves an ion gaining electrons whereas that occurring at the anode involves an atom or ion losing electrons. Metals or hydrogen gas are usually produced at the cathode whereas non-metals or oxygen gas are usually produced at the anode.

26.3 FARADAY'S LAW OF ELECTROLYSIS

This law may be stated as follows:

The mass of substance liberated during electrolysis is proportional to the amount of electric charge that has passed round the circuit

Historically, this result was important because it suggested that a link existed between atoms and electricity, as it established that a certain quantity of electric charge can be associated with a given amount of matter. In the light of our present knowledge of the atom, the law can be explained as follows.

In any electrolysis each type of ion needs to gain or lose a certain number of electrons to cause it to be discharged. Thus, when Cu^{2+} ions are discharged at the cathode *two* electrons need to be gained by *each* ion in order to form a neutral copper atom. If the mass of copper deposited at the cathode is to be doubled, it will require twice as many atoms of copper to be formed from the ions. This will require twice as many electrons to flow onto the cathode. As each electron possesses a fixed amount of electric charge, the total amount of charge that passes round the circuit must be doubled. The total amount of charge that flows round the circuit can be calculated.

As an example of the above ideas, consider the following example. Suppose a current of 1.5 A was recorded on the ammeter in Fig. 26.2 for 50 min. A current of 1.5 A means 1.5 C of charge pass through the meter

in 1 s. Thus, in 50 min (3000 s), the total amount of charge that passes round the circuit is $(3000 \times 1.5) \text{ C} = 4500 \text{ C}$. As each electron possesses a charge of $1.6 \times 10^{-19} \text{ C}$, the number of electrons that pass onto the cathode is

$$\frac{4500}{1.6 \times 10^{-19}} = 2.8 \times 10^{22}$$

Each copper ion requires *two* electrons to discharge it and thus a total of 1.4×10^{22} atoms of copper will be liberated at the cathode. One mole of copper (6×10^{23} atoms of copper) has a mass of 63.5 g and the mass of copper produced at the cathode is

$$\frac{63.5}{6 \times 10^{23}} \times 1.4 \times 10^{22} = 1.48 \text{ g}$$

The above calculation is rather involved and, in order to simplify it, it is convenient to define a quantity called the *electrochemical equivalent* of a substance.

The electrochemical equivalent of a substance is the mass of substance which is liberated during electrolysis when one coulomb of electric charge passes round the circuit

Using this quantity the mass of substance liberated during electrolysis can be calculated using only the number of coulombs that have passed round the circuit.

The electrochemical equivalent of copper is 0.000 33 g/C and this means that 0.000 33 g of copper will be liberated by 1 C of electric charge. Thus, if 4500 C of charge pass round the circuit the mass of copper produced is

$$4500 \times 0.000\ 33 = 1.48 \text{ g}$$

In general, we may write:

$$m = zIt$$

where m is the mass in grams of substance deposited, I the current in amps, t the time in seconds for which it flows, and z the electrochemical equivalent of the substance in grams/coulomb.

26.4 USES OF ELECTROLYSIS

(a) Extraction and purification of metals

Electrolysis is used in the extraction of highly reactive metals from their compounds. The molten compound is used as the electrolyte and the positively charged metal ions are attracted to the cathode where they are

discharged. In general, the molten metal formed at the cathode is more dense than the compound and sinks to the bottom of the tank containing the electrolyte, from where it is extracted.

It is also possible to purify some metals by electrolysis. In order to do this the impure metal acts as the anode and the cathode is made from small samples of the pure metal. The impure metal, and not the impurities, go into solution at the anode, travel through the electrolyte to the cathode where they are discharged and form a coating over the cathode.

(b) Electroplating

Plating is used to protect metals such as steel from corrosion by covering them with a thin layer of a less-reactive metal such as nickel or chromium. The article to be coated is made the cathode and the metal to be used for the coating acts as the anode. Thus, in order to chromium-plate steel, chromium is used as the anode and steel as the cathode. The electrolyte is a chromium salt. The chromium goes into solution at the anode, passes through the electrolyte and is discharged and deposited as a thin layer on the steel.

In silver-plating the article to be silver-plated is made the cathode, the anode is silver and the electrolyte is a silver salt.

26.5 SOURCES OF ELECTRICAL ENERGY

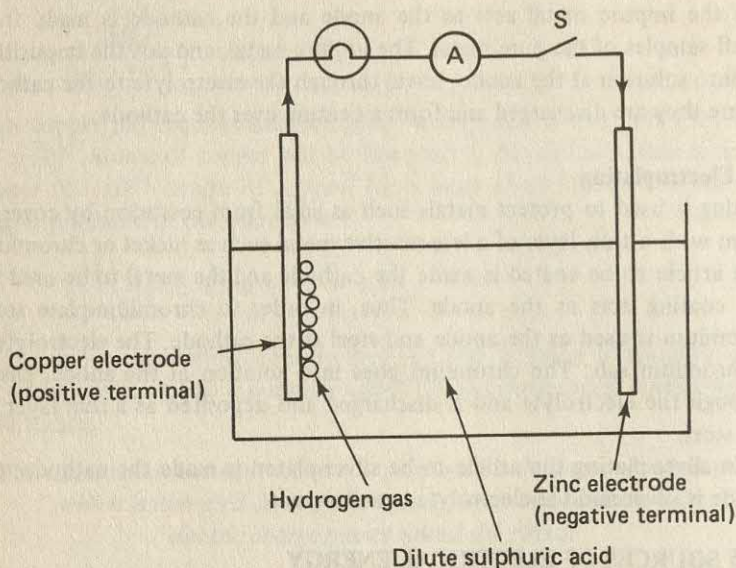
The two most common sources of electrical energy are the mains supply and the battery. As was discussed in Section 25.3, the mains supply uses an alternating voltage and its method of production and distribution will be discussed in Chapter 27.

Batteries on the other hand supply a voltage that is fixed in direction and for this reason the voltage is said to be direct. The energy obtained from a battery comes from the chemical reactions that occur within it, and a number of different batteries (cells) will now be discussed.

26.6 THE SIMPLE CELL

One of the first types of cell used to produce electrical energy is shown in Fig. 26.4. It consists of two electrodes of zinc and copper immersed in dilute sulphuric acid and is traditionally called the *simple cell*. It produces an e.m.f. of about 1 V. When the switch S in Fig. 26.4 is closed, a current is recorded on the ammeter and the bulb lights up. However, the bulb does not stay alight for long. This is because of the production of hydrogen gas, which forms a layer covering the copper, and causes the current in the circuit to drop. This effect, known as *polarisation*, makes the cell of little practical value, as current is obtained for only a short period of time.

fig 26.4 a simple cell. Hydrogen gas formed round the copper prevents the cell supplying current for any prolonged period of time



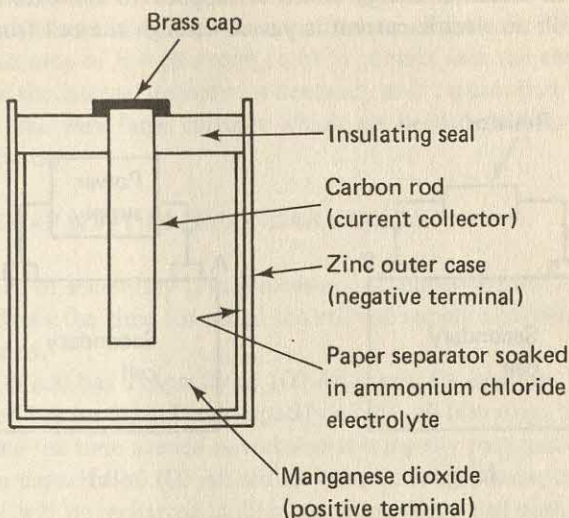
The cell also suffers from another defect in that pure zinc must be used. If pure zinc is not used, hydrogen gas is formed around the zinc itself and the electrons do not flow around the external circuit. This effect is known as *local action* and can be avoided by rubbing mercury over the zinc which prevents the formation of hydrogen on the zinc, and causes the electrons to pass round the external circuit.

26.7 THE DRY (LECLANCHÉ) CELL

This is the most common form of cell now in use and a diagram of the cell is shown in Fig. 26.5. The zinc outer case acts as the negative terminal and the carbon rod is the positive terminal.

The electrolyte is a solution of ammonium chloride soaked into a paper separator. The carbon rod has manganese dioxide powder packed tightly around it. The manganese dioxide is the positive electrode (the carbon rod is simply a current collector), and it also serves to prevent any polarising action which may occur in the cell. This type of cell is used extensively in torches, radios and cassette recorders.

fig 26.5 a dry cell. This is the most common form of cell, and is used in torches, radios and cassette recorders



26.8 THE E.M.F. AND INTERNAL RESISTANCE OF CELLS

The chemical changes which occur within a cell release energy which is supplied to the external circuit. The size of the e.m.f. produced by a cell depends on the nature of the chemical changes which occur in the cell. The e.m.f. of the Leclanché cell is 1.5 V, and that of the simple cell is about 1 V, albeit for a short period of time.

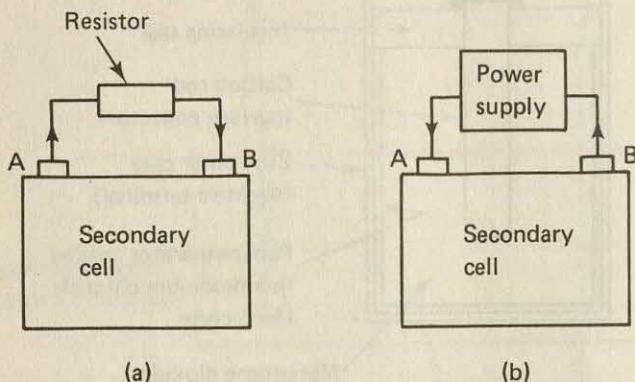
The internal resistance of the cell (see Section 23.8) depends on its construction and on the strength of the electrolyte. The internal resistance is lowered by using terminals which have a large surface area in contact with the electrolyte and which are placed close together.

26.9 PRIMARY AND SECONDARY CELLS

All the cells so far discussed are called *primary* cells and the chemical changes which occur within them cannot be reversed. Thus, when all the chemicals in the cell have been used up, no more energy can be obtained from the cell and it is thrown away.

However, it is possible to obtain cells in which the chemical changes can be reversed and such cells are called *secondary* cells. This reversal in the reaction is brought about by passing a current through the cell in the opposite direction to that in which it flows when the cell supplies energy.

These ideas can be best understood by considering Fig. 26.6a and b. In Fig. 26.6a the cell is being used as a source of energy and the current flows from terminal A to terminal B. The chemical energy in the cell is being converted to electrical energy which is supplied to the external circuit. In Fig. 26.6b an electric current is passed through the cell from terminal B to terminal A. This process is known as *recharging*.



A to terminal B by connecting it to a power supply which has a greater e.m.f. than the cell. The electrical energy from the supply is used to reform the chemicals within the cell which were used up during the discharge. This process is known as *recharging*.

26.10 THE LEAD-ACID ACCUMULATOR

The lead-acid accumulator, which is the type of accumulator used as a car battery, consists of six secondary cells connected in series. As each cell produces an e.m.f. of about 2.0 V, the total e.m.f. produced by the accumulator is about 12 V.

Each secondary cell consists of two electrodes immersed in sulphuric acid. The negative electrode is made of lead and the positive electrode consists of a lead plate coated with lead dioxide. When used as a source of energy both plates gradually become coated with lead sulphate and the concentration of the sulphuric acid drops. When the cell is recharged a large amount of the lead sulphate is converted back to lead and lead dioxide.

As the concentration of the sulphuric acid falls when the cell supplies energy, the specific gravity (density) of the acid is used in order to determine its state of charge. When the specific gravity falls below 1.18, the cell should be recharged. The cell can be seriously damaged if the electrodes are not completely covered with electrolyte and it is important to 'top up'

the cell from time to time with distilled water, to replace that lost by evaporation. Apart from this the cell requires no maintenance.

The lead-acid cell is particularly useful for a car battery as it has very low internal resistance. This results from the low resistance of the sulphuric acid electrolyte and also from the construction of the cell which allows a large surface area of the electrode to be in contact with the electrolyte. A low value of the internal resistance is necessary as it ensures that the battery can supply the very large currents which are needed to turn the starter motor of the car.

26.11 THE CAPACITY OF SECONDARY CELLS

The capacity of secondary cells is measured in ampere-hours (Ah) and this quantity allows the time for which the cell will supply a certain current to be determined.

Thus, if a cell has a capacity of 100 Ah it will, for example, supply 1 A for 100 h or 5 A for 20 h. If the capacity of the cell is known, it is possible to determine the time needed to recharge it using any particular current. If a cell has a capacity of 100 Ah and is being recharged using a current of 2A, then it will be recharged in 50 h, assuming that it had been allowed to run down completely.

WORKED EXAMPLE

A copper voltameter is in series with a silver voltameter and a resistor of resistance $2\ \Omega$. 0.4 g of copper are deposited in 20 min. (i) What is the mass of silver deposited in 20 min? (ii) What is the p.d. across the resistor? (e.c.e. of copper is 0.00033 g/C and e.c.e. of silver is 0.00112 g/C).

- (i) $m = zIt$ (page 276). Since the same current flows in each voltameter for the same time the masses deposited are proportional to the electrochemical equivalents.

$$\begin{aligned}\text{Hence, mass of silver deposited} &= \frac{0.4}{0.00033} \times 0.00112 \\ &= 1.36\text{ g}\end{aligned}$$

(ii) $m = zIt$ $0.4 = 0.00033 \times I \times (20 \times 60)$

$$I = 1.01\text{ A}$$

$$\text{p.d. across } 2\ \Omega \text{ resistor} = 1.01 \times 2 = 2.02\text{ V}$$

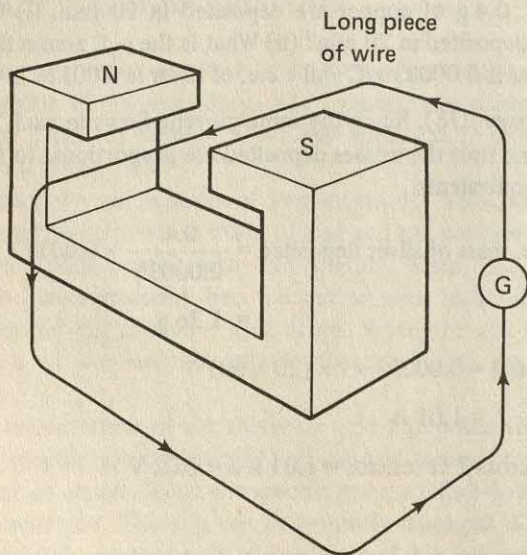
ELECTROMAGNETIC INDUCTION

27.1 INTRODUCTION

Electromagnetic induction is the term used to describe the effect by which a voltage or e.m.f. can be induced in a conductor.

Fig. 27.1 shows a simple experimental arrangement to demonstrate this phenomenon. It consists of a long piece of wire which is held between the poles of a permanent magnet and connected in series to a very sensitive galvanometer G.

fig 27.1 *if either the wire or the magnet is moved up or down a deflection will be observed on the galvanometer indicating that an e.m.f. has been induced in the wire*

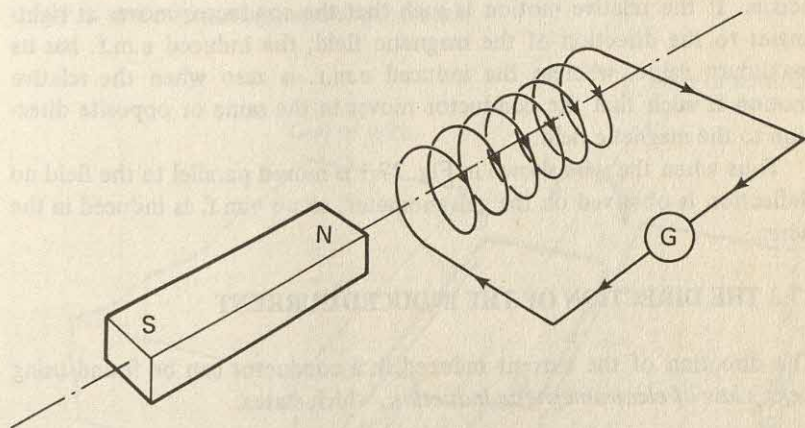


When the wire is held stationary in the magnetic field no deflection is observed on the galvanometer. However, when the wire is moved down a current is recorded on the galvanometer. In order to produce this current an e.m.f. must have been induced in the wire. If the wire is now moved up the galvanometer again shows a deflection, but in this case in the opposite direction. This shows that the direction of the e.m.f. induced in the wire depends on the direction in which the wire is moved.

An e.m.f. can also be induced in the wire if the wire itself remains fixed in position and the magnet is moved up or down. Again the direction of the deflection indicates that the direction of the induced e.m.f. depends on the direction in which the magnet is moved.

Similar effects to this can be observed using the apparatus shown in Fig. 27.2. This consists of a galvanometer G connected across the ends of a

fig 27.2 *when the magnet and the solenoid are being moved further apart, the deflection on the galvanometer is opposite to that which occurs when they are being moved closer together*



solenoid which is positioned close to a bar magnet. If either the magnet or the solenoid is moved, a current is recorded on the galvanometer, and in order to produce this current an e.m.f. must have been induced in the solenoid. However, no deflection is observed when the magnet and solenoid are stationary or moved with the same velocity. It is also found that when the solenoid and magnet are being moved closer together the deflection on the galvanometer is in the opposite direction to that which occurs when they are being moved further apart.

The results of both these experiments may be summarised as follows:

Relative motion between a conductor and a magnetic field is necessary if an e.m.f. is to be induced in the conductor

The direction of the induced e.m.f. is reversed if the relative motion between the magnetic field and the conductor is reversed

27.2 THE SIZE OF THE INDUCED E.M.F.

In both the experiments described above it is found that the faster the conductor or the magnet is moved the larger the deflection recorded on the galvanometer. This indicates that the size of the e.m.f. induced in a conductor depends on the speed of the relative motion between the conductor and the magnetic field. This result is known as *Faraday's law of electromagnetic induction* and may be stated as follows:

Whenever an e.m.f. is induced in a conductor due to the relative motion of the conductor and a magnetic field, the size of the induced e.m.f. is proportional to the speed of the relative motion

The size of the induced e.m.f. is also found to depend on the strength of the magnetic field and on the direction in which the relative motion occurs. If the relative motion is such that the conductor moves at right-angles to the direction of the magnetic field, the induced e.m.f. has its maximum value, whereas the induced e.m.f. is zero when the relative motion is such that the conductor moves in the same or opposite direction to the magnetic field.

Thus when the wire shown in Fig. 27.1 is moved parallel to the field no deflection is observed on the galvanometer, so no e.m.f. is induced in the wire.

27.3 THE DIRECTION OF THE INDUCED CURRENT

The direction of the current induced in a conductor can be found using *Lenz's law of electromagnetic induction*, which states:

The direction of any current induced in a conductor is such that it opposes the change or motion producing it

As the first example of the use of this law, consider the situation when the wire in Fig. 27.1 is moved down. This downward movement of the wire causes a current to be induced in it. This current must flow in such a direction as to produce an upward force on the wire, as only in this way can the downward movement of the wire be opposed. The direction in which the induced current must flow in order to produce an upward force is found by using the left-hand rule (Section 24.5(c)) and the direction of the current is indicated by the arrows shown in Fig. 27.1.

As a second example of the use of this law, consider the situation when the magnet in Fig. 27.2 is moved away from the solenoid. This will cause

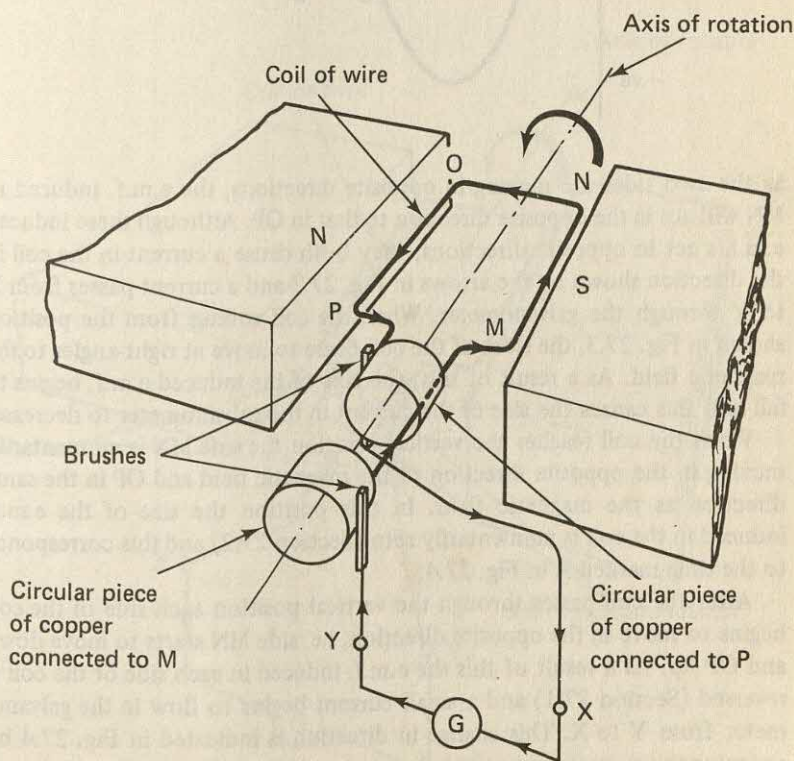
an induced current in the solenoid, which according to Lenz's law must be in such a direction as to oppose the movement of the magnet. This movement will be opposed if the end of the solenoid closest to the magnet acts as a *S-pole*. Using the right-hand rule (Section 24.1), this occurs when the current in the solenoid is in the direction shown in Fig. 27.2.

27.4 THE ALTERNATING-CURRENT GENERATOR

The fact that an e.m.f. can be induced in a conductor when it is moved in a magnetic field is made use of in the alternating-current generator.

Fig. 27.3 shows a simple form of such a generator. It consists of a coil of wire which can be rotated in a magnetic field. Each end of the coil is attached to a circular piece of copper (called a slip ring), which is fixed to, but insulated from, the shaft on which the coil is wound. (This shaft is not shown in Fig. 27.3.) Each terminal of the generator, X and Y, is connected to one of these pieces of copper by means of a brush.

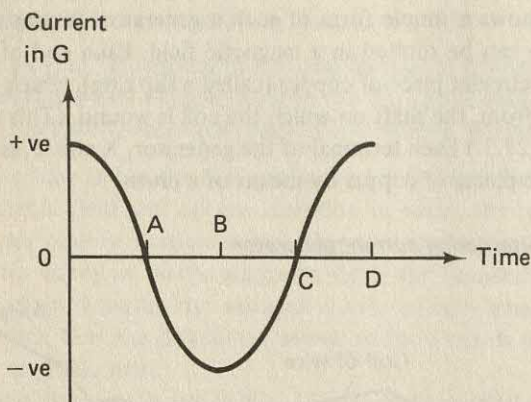
fig 27.3 *the alternating-current generator*



When the coil is rotated in the magnetic field, an alternating voltage is produced across the terminals X and Y and this causes an alternating current in the galvanometer G. The variation of this current with time for one complete rotation of the coil is shown in Fig. 27.4.

At the time marked O in Fig. 27.4, the coil is in the position shown in Fig. 27.3. In this position both sides of the coil, MN and OP, are momentarily moving at right-angles to the magnetic field. As discussed in Section 27.2, this causes a large e.m.f. to be induced in each side of the coil; and

fig 27.4



as the two sides are moving in opposite directions, the e.m.f. induced in MN will act in the opposite direction to that in OP. Although these induced e.m.f.'s act in opposite directions, they both cause a current in the coil in the direction shown by the arrows in Fig. 27.3 and a current passes from X to Y through the galvanometer. When the coil rotates from the position shown in Fig. 27.3, the sides of the coil cease to move at right-angles to the magnetic field. As a result of this, the size of the induced e.m.f. begins to fall and this causes the size of the current in the galvanometer to decrease.

When the coil reaches the vertical position the side MN is momentarily moving in the opposite direction to the magnetic field and OP in the same direction as the magnetic field. In this position the size of the e.m.f. induced in the coil is momentarily zero (Section 27.2) and this corresponds to the time marked A in Fig. 27.4.

After the coil passes through the vertical position each side of the coil begins to move in the opposite direction, i.e. side MN starts to move down and OP up. As a result of this the e.m.f. induced in each side of the coil is reversed (Section 27.1) and a small current begins to flow in the galvanometer from Y to X. This change in direction is indicated in Fig. 27.4 by assigning the current a negative value.

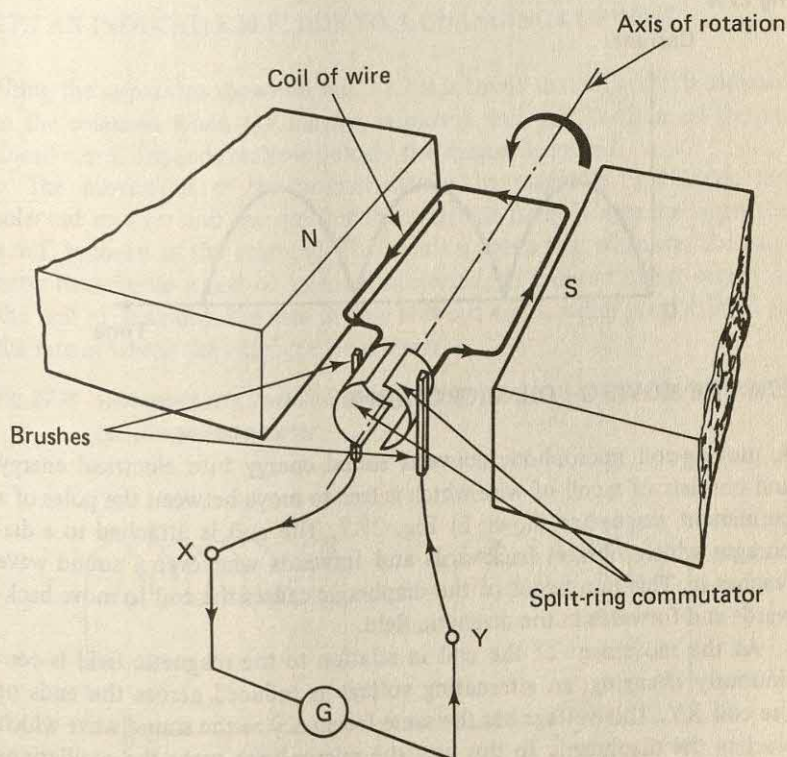
At the time marked B in Fig. 27.4 the coil is again in the horizontal position and because each side of the coil is moving at right-angles to the field the current in the galvanometer has its maximum value.

As the coil continues to rotate the current will again drop to zero, when the coil passes through the vertical position, and finally the coil will return to the position shown in Fig. 27.3 at the time marked D in Fig. 27.4.

Thus, as the coil rotates, each side of the coil moves in a continually changing direction in relation to the magnetic field and this produces an alternating current in the galvanometer. The frequency of this current is the same as the frequency of rotation of the coil.

In commercial generators used at power stations the coil is fixed in position and the magnet is rotated, using steam turbines. Rotating the magnet has the advantage that it avoids the need to use brushes, which are required if the coil rotates.

fig 27.5 *the direct-current generator*



27.5 THE DIRECT-CURRENT GENERATOR

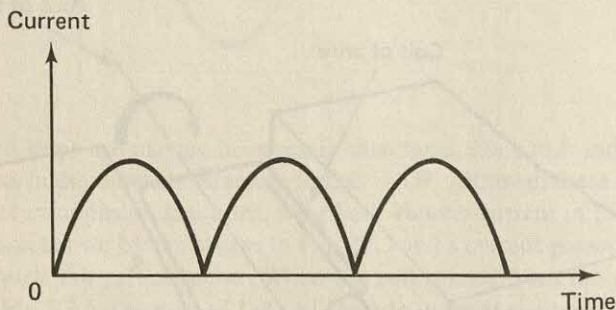
The construction of a direct-current generator, shown in Fig. 27.5, is very similar to that of the alternating-current generator. The only difference is that the electrical connection between the ends of the coil and the terminals X and Y is made using a split-ring commutator.

The split-ring commutator reverses the electrical connection between the terminals of the generator and the ends of the coil each time the coil passes through the vertical position.

The fact that this change is made to occur at exactly the same time as the current in the coil reverses direction means that the current through the galvanometer is always in one direction, and because of this the generator is said to produce a direct current.

The variation in this current with time is shown in Fig. 27.6 and, as can be seen from this graph, the current is by no means steady. A much steadier current can be produced by using a radial magnetic field (Section 24.6) as in this case each side of the coil moves at right-angles to the magnetic field for a much longer time, and because of this a steadier current is produced.

fig 27.6



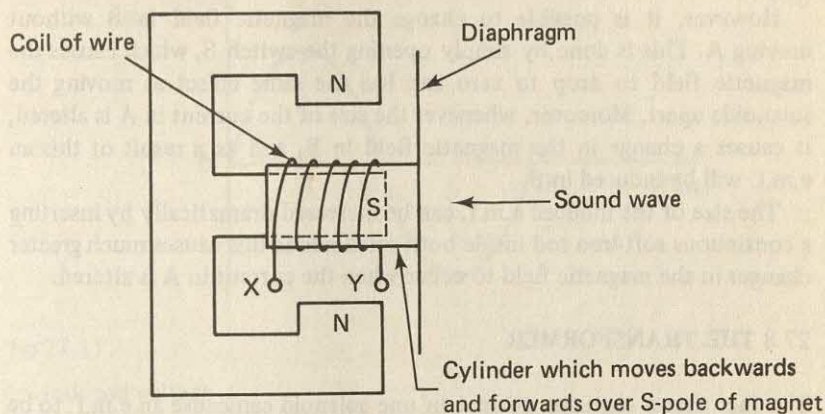
27.6 THE MOVING-COIL MICROPHONE

A moving-coil microphone converts sound energy into electrical energy and consists of a coil of wire which is free to move between the poles of a permanent magnet as shown in Fig. 27.7. The coil is attached to a diaphragm which vibrates backwards and forwards whenever a sound wave reaches it. This movement of the diaphragm causes the coil to move backwards and forwards in the magnetic field.

As the movement of the coil in relation to the magnetic field is continuously changing, an alternating voltage is induced across the ends of the coil XY. This voltage has the same frequency as the sound wave which reaches the diaphragm. In this way the microphone turns the oscillations

caused by the sound wave impinging on the diaphragm into an alternating voltage of the same frequency as the sound wave.

fig 27.7 *the moving-coil microphone*

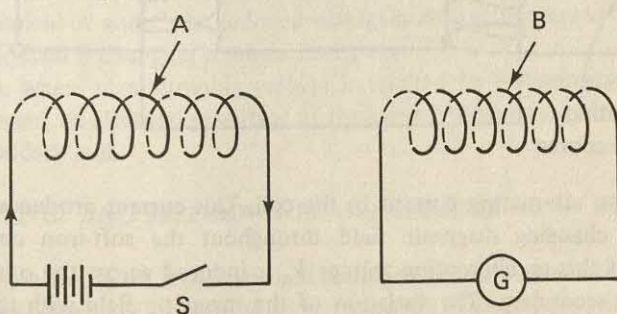


27.7 AN INDUCED E.M.F. DUE TO A CHANGING CURRENT

Using the apparatus shown in Fig. 27.2 it is found that an e.m.f. is induced in the solenoid when the magnet is moved and that the size of this induced e.m.f. depends on how quickly the magnet is moved.

The movement of the magnet causes the magnetic field inside the solenoid to alter and the quicker the magnetic field changes the larger the e.m.f. induced in the solenoid. This result suggests that whenever the magnetic field inside a coil or solenoid is altered, an induced e.m.f. occurs in the coil or solenoid, the size of this induced e.m.f. being proportional to the rate at which the magnetic field alters.

fig 27.8 *whenever the switch S is opened or closed, a current is recorded on the galvanometer*



In Section 24.2, it was shown that the current in a solenoid produces a magnetic field similar to that of a bar magnet. Thus, in Fig. 27.8, when the switch is closed and solenoid A is moved towards or away from solenoid B an e.m.f. is induced in B, as moving A is then equivalent to moving a bar magnet.

However, it is possible to change the magnetic field in B without moving A. This is done by simply opening the switch S, which causes the magnetic field to drop to zero and has the same effect as moving the solenoids apart. Moreover, whenever the size of the current in A is altered, it causes a change in the magnetic field in B, and as a result of this an e.m.f. will be induced in B.

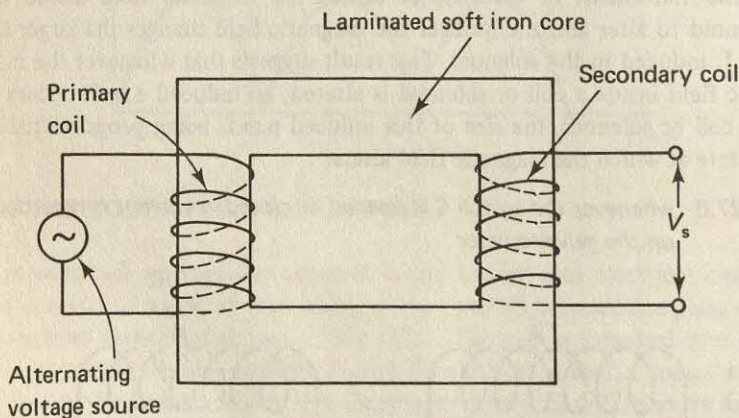
The size of the induced e.m.f. can be increased dramatically by inserting a continuous soft-iron rod inside both solenoids as this causes much greater changes in the magnetic field to occur when the current in A is altered.

27.8 THE TRANSFORMER

The fact that a changing current in one solenoid can cause an e.m.f. to be induced in another solenoid is made use of in the transformer.

Fig. 27.9 shows a diagram of a transformer. It consists simply of two coils of wire wound on a continuous soft-iron core. One of the coils, called the primary, is connected to an alternating voltage source and this

fig 27.9 *the transformer. The alternating current in the primary coil induces an alternating voltage V_s across the secondary coil*



produces an alternating current in the coil. This current produces a continuously changing magnetic field throughout the soft-iron core and because of this an alternating voltage V_s is induced across the other coil, called the secondary. The variation of the magnetic field with time and

of the induced voltage V_s with time are shown in Figs. 27.10 and 27.11 respectively.

fig 27.10

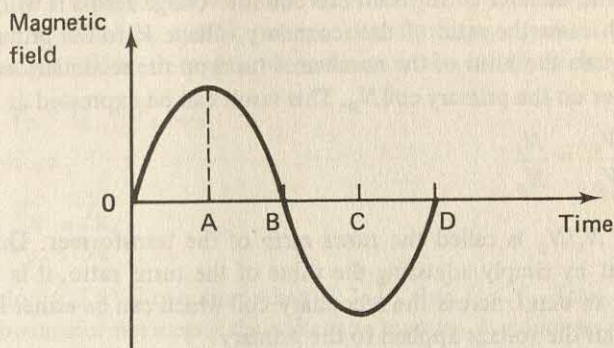
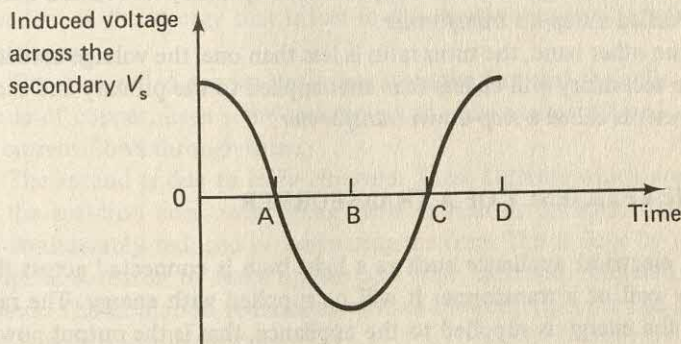


fig 27.11



The size of the induced e.m.f. in *each turn* of the secondary coil depends on the rate at which the magnetic field through it alters and its direction depends on whether the field is increasing or decreasing. At the times marked O and D in Figs. 27.10 and 27.11, the induced voltage has a large positive value as the field is increasing at its maximum rate, whereas at the times marked A and C the induced voltage is zero as the rate at which the magnetic field is changing is momentarily zero.

Thus, when an alternating voltage is applied to the primary coil of a transformer, an alternating voltage of the same frequency is induced across the secondary coil.

27.9 STEP-UP AND STEP-DOWN TRANSFORMERS

If a transformer with a primary coil having 60 turns and a secondary with 120 turns has a 12 V alternating source applied to the primary, the induced

voltage across the secondary is found to be 24 V. If the number of turns on the secondary is doubled the induced voltage increases to 48 V, as *each turn* on the secondary coil has the same e.m.f. induced in it. Thus, by doubling the number of turns on this coil the voltage across it will double.

In both cases the ratio of the secondary voltage V_s to the primary voltage V_p equals the ratio of the number of turns on the secondary coil N_s to the number on the primary coil N_p . This result can be expressed as follows:

$$\frac{V_s}{V_p} = \frac{N_s}{N_p} \quad (27.1)$$

The ratio N_s/N_p is called the *turns ratio* of the transformer. This result shows that by simply adjusting the value of the turns ratio, it is possible to induce an e.m.f. across the secondary coil which can be either larger or smaller than the voltage applied to the primary.

If the turns ratio is greater than one, the induced voltage across the secondary will be larger than that applied to the primary and such a transformer is called a *step-up transformer*.

If, on the other hand, the turns ratio is less than one, the voltage induced across the secondary will be less than that applied to the primary and such a transformer is called a *step-down transformer*.

27.10 THE EFFICIENCY OF A TRANSFORMER

When an electrical appliance such as a light bulb is connected across the secondary coil of a transformer it will be supplied with energy. The rate at which the energy is supplied to the appliance, that is the output power of the transformer, depends on the secondary voltage V_s and the secondary current I_s . Using equation (25.4) the output power is given by the following equation:

$$\text{Output power} = V_s \times I_s \quad (27.2)$$

The energy supplied to the appliance originates from the power supply connected to the primary coil. The rate at which this energy is supplied, that is the input power to the transformer, can again be calculated using equation (25.4) and is given by the following equation:

$$\text{Input power} = V_p \times I_p \quad (27.3)$$

where V_p is the voltage across the primary coil and I_p is the current in the primary coil.

Using these two results the efficiency of the transformer is defined in the following way:

$$\text{Efficiency} = \frac{V_s \times I_s}{V_p \times I_p} \times 100\% \quad (27.4)$$

Thus, if a transformer is 100% efficient (most transformers have an efficiency approaching 100%), it means that all the energy from the power supply is delivered to the appliance and in this case it follows that:

$$V_p \times I_p = V_s \times I_s \quad (27.5)$$

and therefore

$$\frac{V_s}{V_p} = \frac{I_p}{I_s} \quad (27.6)$$

Equation (27.6) shows that when $V_s > V_p$ then $I_s < I_p$. Thus, although a step-up transformer causes the voltage to increase, it causes an associated decrease in the current.

Most transformers have an efficiency greater than 95%, and the small fraction of the energy that is lost in the transformer itself arises for two main reasons.

The first, called *copper loss*, arises from the fact that the coils, although made of copper, have some resistance and some energy will be lost as heat as current flows through them.

The second is due to *eddy currents*. These currents which are induced in the soft-iron core cause energy loss as heat in the core. This loss can be considerably reduced by laminating the core. This is done by using thin strips of soft iron to make up the core, with each strip insulated from the others. The insulation reduces the size of the eddy currents and as a result less energy is lost in the core.

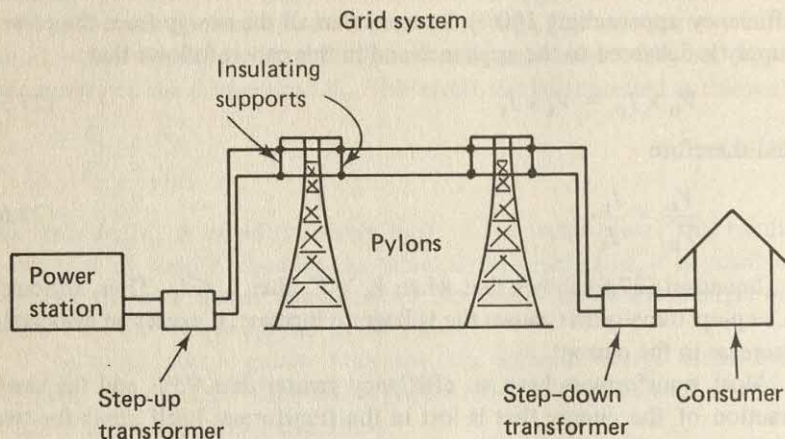
27.11 THE DISTRIBUTION OF ELECTRICAL ENERGY

The electrical energy used in farms, factories and houses is delivered through a system of cables, called the *National Grid*, which links the consumer to the power stations throughout the country. These cables are made of thick copper wire and are mostly supported on pylons.

Although the mains supply is delivered to houses at 240 V, it is transmitted over much of the grid system at voltages of 275 000 V or 400 000 V, and step-down transformers are used to convert this voltage to the much lower values used by consumers.

The high voltages used in transmission are not generated at the power station but are produced by connecting step-up transformers between the grid system and the power station. Fig. 27.12 shows the essential features of the system used to transmit the energy from the power station to the consumer.

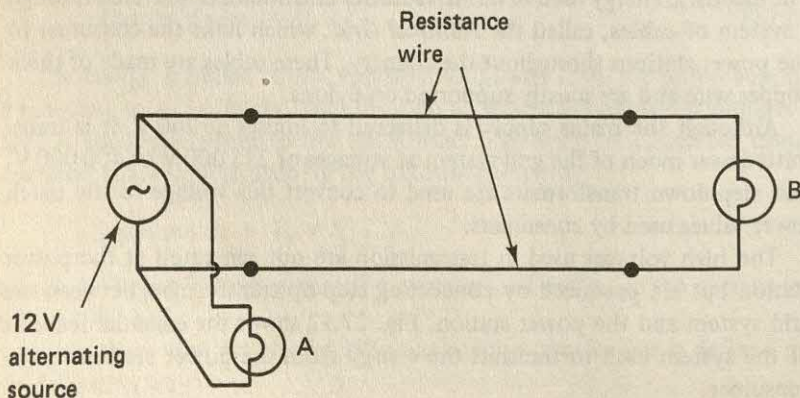
fig 27.12 *the energy from the power station is delivered to the consumer using a system of overhead cables called the National Grid system*



At first sight it appears rather strange to step up the voltage for transmission and then to step it down for use by the consumer. The reason for doing this can be demonstrated using the apparatus shown in Fig. 27.13.

This consists of a 12 V alternating supply which has two similar bulbs, A and B, connected across its terminals. One of the bulbs, A, is connected to the supply using short leads whilst the other, B, is connected to the supply via two pieces of resistance wire. These comparatively short lengths of resistance wire behave in a similar manner to the much longer lengths of copper wire which would be used if the electricity were to be distributed over much greater distances.

fig 27.13



When the power supply is switched on, it is found that the brightness of bulb A is much greater than that of B. Clearly, bulb B is being supplied with energy at a much lower rate than A, and a lot of the energy which should be supplied to B is lost as heat in the resistance wires.

The power supply is now switched off and a step-up transformer having a turns ratio of 20:1 is inserted between the supply and the resistance wires. A step-down transformer of turns ratio 1:20 is also inserted between the wires and the bulb B. With the transformers inserted the arrangement becomes similar to that shown in Fig. 27.12.

When the power supply is now switched on it is found that both bulbs have a similar brightness which indicates that the energy lost in the resistance wires has been greatly reduced. The reason that this occurs can be explained as follows.

The step-up transformer has increased the voltage across the resistance wires to 240 V as it has a turns ratio of 20:1 (see equation (27.1)). Assuming that the transformer is 100% efficient this increase in the voltage across the resistance wires must be associated with a corresponding drop in the current. The current in the wires will be reduced by a factor of 20 (see equation (27.6)). As a result of this much smaller current in the resistance wires, the energy lost in them as heat is greatly reduced.

Thus, by stepping-up the voltage at which the electricity is transmitted it is possible to reduce the energy loss in the connecting wires, and it is precisely for this reason that such high voltages are used in much of the grid system.

Finally, it must be pointed out that such a system is used only because of the very high efficiency of the transformer. If the transformer itself did not have such a high efficiency then the energy loss in the transformer could be greater than that which occurred in the transmission lines, in which case using the transformers would be counter-productive.



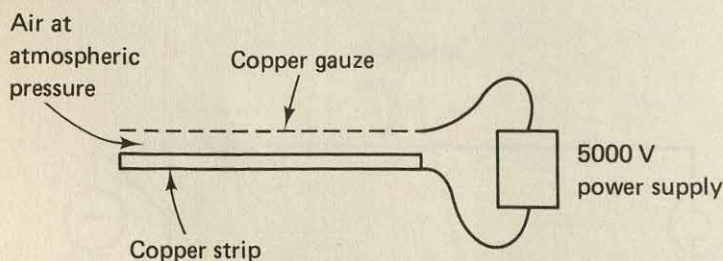
RADIOACTIVITY AND THE ATOM

28.1 THE SPARK COUNTER

Radioactive substances have the ability to ionise the air surrounding them and it is this property that is used to distinguish them from materials which are not radioactive.

The simplest experiment that shows this uses a spark counter (Fig. 28.1). This consists of a piece of copper gauze supported above a copper strip. The distance between the gauze and strip is just long enough to prevent a spark when a potential difference of 5000 V is applied across them. However, when certain substances are placed close to the gauze sparks are seen to occur between the gauze and the strip. As was discussed in Section 22.8, sparks can only occur when the air between the gauze and strip is ionised and materials which emit radiation to cause this ionisation are said to be radioactive.

fig 28.1 *a spark counter. When the air between the gauze and strip is ionised, a spark occurs*

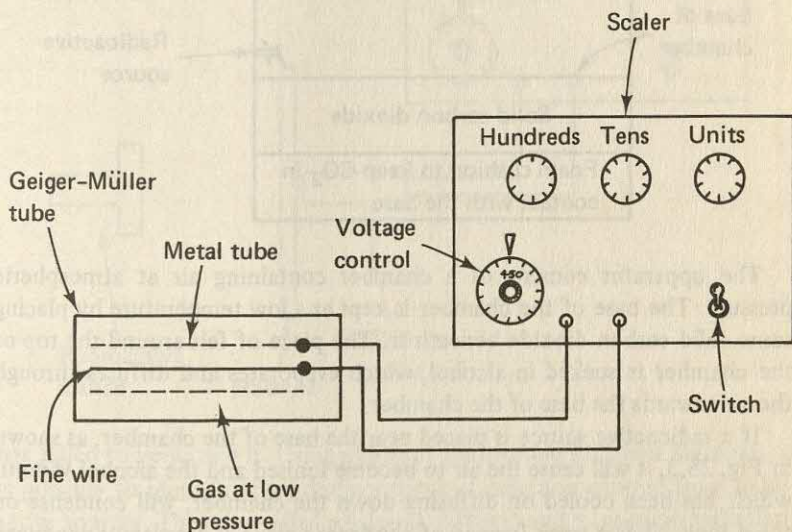


28.2 THE GEIGER-MÜLLER TUBE

The spark counter requires a high level of ionisation to produce a visible spark, and in order to detect radioactive sources which produce lower levels of ionisation a Geiger-Müller (GM) tube is used.

This consists of a cylindrical metal tube with a fine wire running along the axis of the cylinder (Fig. 28.2). The tube is filled with an inert gas such as neon at a low pressure and a potential difference is applied between the cylinder and the wire. The potential difference is usually applied by connecting the tube to a *scaler* or *ratemeter* and a control on the front panel of the instruments records and the potential difference applied to the tube. This potential difference is usually of the order of 450 V.

fig 28.2 a Geiger-Müller tube and scaler



When the tube, connected to a scaler, is placed near a radioactive source, the neon gas is ionised and a pulse of current will occur in the tube. These pulses of current are automatically counted and recorded by the scaler, and each pulse corresponds to one ionising radiation passing through the tube.

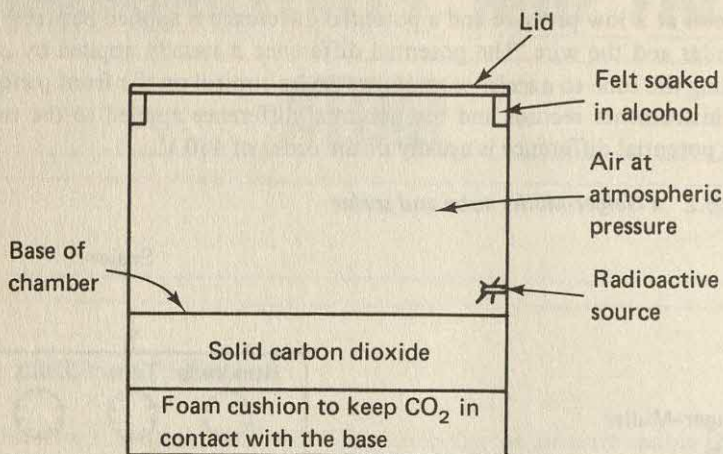
Both the current pulses in the GM tube and the sparks in the spark counter occur at random, and this indicates that the radiations from the source which produce ionisation are not emitted in a regular way. Thus:

The radiations which cause ionisation are emitted from a radioactive material in a random way

28.3 THE DIFFUSION CLOUD CHAMBER

A diagram of a diffusion cloud chamber is shown in Fig. 28.3, and this method of detecting radioactive substances again uses their ability to cause ionisation.

fig 28.3 a diffusion cloud chamber



The apparatus consists of a chamber containing air at atmospheric pressure. The base of the chamber is kept at a low temperature by placing some solid carbon dioxide beneath it. The piece of felt around the top of the chamber is soaked in alcohol, which evaporates and diffuses through the air towards the base of the chamber.

If a radioactive source is placed near the base of the chamber, as shown in Fig. 28.3, it will cause the air to become ionised and the alcohol vapour, which has been cooled on diffusing down the chamber, will condense on these ions. Thus a track formed of condensed alcohol droplets is produced along the path of the ionising radiation emitted from the source. This track is only visible for a short time before the liquid drops of alcohol fall to the base of the chamber.

The tracks are observed to occur at random and this again supports the view that the ionising radiations are emitted from the source in a random manner.

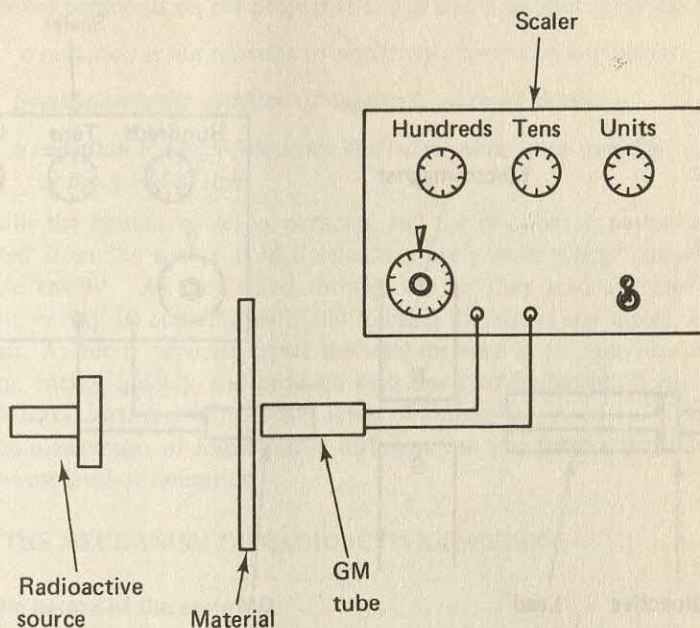
28.4 THE PROPERTIES OF THE IONISING RADIATIONS EMITTED FROM RADIOACTIVE SUBSTANCES

(a) Penetrating powers and range in air

The penetrating powers of the radiations which cause ionisation can be investigated using the apparatus shown in Fig. 28.4.

A GM tube connected to a scaler is placed close to a radioactive source and various materials are inserted between the source and the tube. If the number of current pulses recorded on the scaler is reduced when the material is inserted, it means that some of the radiation from the source

fig 28.4 *apparatus to test the penetrating powers of the radiations emitted from radioactive sources*



has failed to penetrate it. Using different sources and different materials, it is possible to identify three distinct types of radiation emitted from radioactive sources. These are called α (alpha), β (beta) and γ (gamma) radiation.

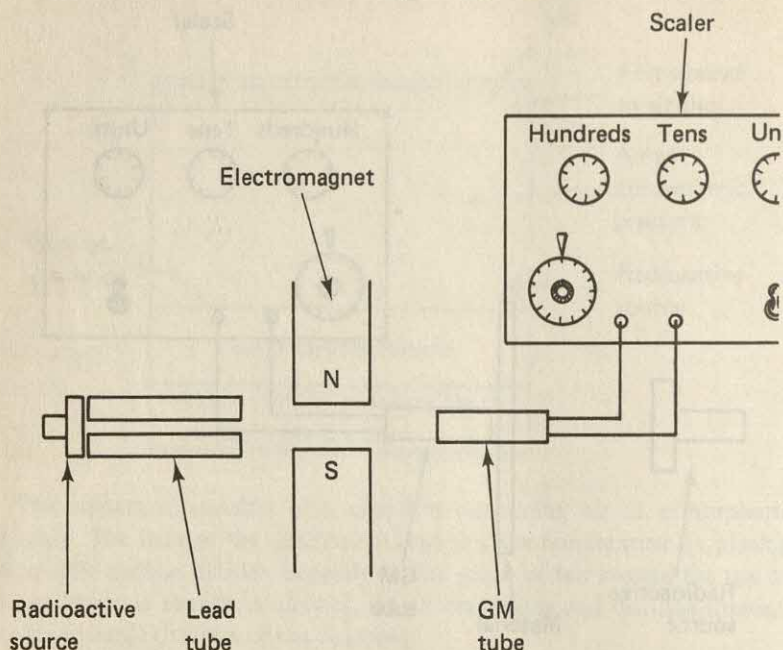
α radiation is unable to penetrate a sheet of paper, whereas β radiation requires a few millimetres thickness of aluminium to stop it. For γ radiation a few centimetres of lead are required to produce any change in the count rate recorded on the scaler. Thus γ radiation is most penetrating and α the least.

The range of the various radiations in air can be investigated by simply moving a GM tube further from the source and noting any change in the count rate recorded on the scaler. If this is done it is found that α radiation, being the least penetrating, will travel less than 10 cm in air, whereas β radiation will travel up to several metres. For γ radiation the reading on the scale gets progressively smaller as the distance between the source and the GM tube increases. However, it is found that γ radiation can be detected at very much greater distances from the source than either α or β radiation.

(b) Deflection by a magnetic field

The apparatus shown in Fig. 28.5 can be used to investigate the effect of a magnetic field on the ionising radiations emitted from radioactive sources.

fig 28.5 *apparatus to test whether the radiations emitted from a radioactive source are deflected by a magnetic field*



A narrow beam of radiation is formed by placing the source at one end of a lead tube. This beam then passes between the poles of an electromagnet and can be detected using a GM tube and scaler. If the count rate recorded on the scaler is reduced when the electromagnet is switched on, it indicates that some of the radiation has been deflected by the magnetic field.

Using this arrangement it is possible to show that γ radiation is not deflected by a magnetic field, and in this respect γ radiation behaves in the same way as a beam of light.

β radiation is deflected, and the direction of the deflection can be found by moving the GM tube until it records a high count rate with the electromagnet switched *on*. The direction of the deflection is the same as would occur if a stream of negatively charged particles passed between the poles of the magnet. This similarity in behaviour suggests that β radiation possesses a negative charge.

As α radiation has such a short range in air it is not possible to use the apparatus in Fig. 28.5 to determine its behaviour in a magnetic field. However, using different experimental techniques it is possible to show that α radiation is deflected by a magnetic field and the direction of the deflection indicates that α radiation possesses a positive charge.

28.5 THE NATURE OF α , β AND γ RADIATION

Further experiments on the properties of α , β and γ radiation indicate that:

α radiation is the emission of positively charged helium nuclei

β radiation is the emission of negatively charged electrons

γ radiation is a high-frequency electromagnetic wave travelling at the speed of light

Both the helium nuclei, α particles, and the electrons, β particles, are emitted from the source at high velocities and possess a large amount of kinetic energy. As they travel through the air they gradually lose this kinetic energy in colliding with and ionising the atoms and molecules in the air. As the α particles travel the least distance in air, they lose their kinetic energy quickly and produce high levels of ionisation. β particles which travel further produce lower levels of ionisation.

The mechanism of ionisation is different for γ radiation; it produces the lowest level of ionisation.

28.6 THE MECHANISM OF RADIOACTIVE EMISSIONS

(a) The nature of the atom

Before discussing how α , β and γ radiations are emitted from a radioactive source it is necessary to consider in more detail the nature of the atom itself.

As was discussed in Section 22.3 atoms consist of protons, neutrons and electrons. The protons and neutrons make up the *nucleus* of the atom and this is positively charged due to the positive charge on the protons. The electrons which have a negative charge surround the nucleus in various orbits. An atom is uncharged and so the number of electrons equals the number of protons. As the mass of both a proton and a neutron is about 1800 times larger than that of an electron, the vast majority of the mass of any atom is contained in its nucleus and the nucleus itself occupies a very small part of the volume occupied by the atom. Thus:

An atom consists of a very small massive positively charged nucleus surrounded by negatively charged orbital electrons

This model of the atom was first proposed by Rutherford as a result of an experiment performed by Geiger and Marsden. In this experiment a beam of α particles from a radioactive source was directed onto a thin piece of gold foil and the resulting deflection of the α particles was recorded. Whilst the vast majority of the α particles were found to pass through the foil undeflected, a very small number were found to be deflected through large angles, and some through angles greater than 90° .

To account for these results Rutherford proposed that all the positive charge in an atom and most of its mass was concentrated into a *very small volume within the atom* called the nucleus. Using this model only a very small number of the positively charged α particles directed at the foil will pass sufficiently close to the positively charged nuclei of the gold atoms to be deflected by them, and as a result the vast majority of the α particles pass straight through the large spaces between the nuclei of the gold atoms.

Using this model of the atom the following quantities can be defined:

The atomic number A of an atom equals the number of protons in its nucleus

The mass number M of an atom equals the sum of the number of protons and neutrons in the nucleus of the atom

All atoms of a given element have the same number of protons in the nucleus and hence contain the same number of orbiting electrons. However, atoms of a given element do vary in the number of neutrons in the nucleus.

Isotopes are atoms of the same element which differ only in the number of neutrons in the nucleus.

It is convenient to represent the mass number M and atomic number A of an atom whose chemical symbol is X in the following manner ${}^M_A X$. Thus, ${}^{238}_{92}\text{U}$ and ${}^{235}_{92}\text{U}$ represent two isotopes of uranium. Both have 92 protons in the nucleus, but the ${}^{238}_{92}\text{U}$ isotope has 146 (i.e. $238 - 92$) neutrons in the nucleus whereas the ${}^{235}_{92}\text{U}$ isotope has only 143 (i.e. $235 - 92$) neutrons in the nucleus.

(b) Nuclear changes resulting from the emission of α , β and γ radiation

α , β and γ radiations are emitted from the nucleus of an atom

These emissions occur because the nucleus of a radioactive atom is unstable, and the change in the nucleus resulting from the emission of each type of radiation will now be considered.

(i) α emission

The helium nucleus, or α particle, consists of two protons and two neutrons and, as a result of this, when an α particle is emitted from a nucleus it causes both the number of protons and the number of neutrons in the nucleus to fall by two. This results in the atomic number falling by two and the mass number by four. Thus, the isotope ${}^M_A X$ changes to the isotope ${}^{M-4}_{A-2} Y$ when it emits an α particle.

(ii) β emission

The emission of a β particle results in the atomic number increasing by one

and the mass number remaining unchanged. Thus, the isotope ${}^M_A X$ is changed to the isotope ${}^{M}_{A+1} Z$ by the emission of a β particle. For this change to occur it is envisaged that one of the neutrons in the nucleus is converted to a proton and an electron. The electron is emitted as the β particle and as the nucleus now contains an extra proton the atomic number has increased by one.

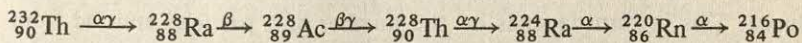
(iii) γ emission

The emission of a γ ray causes no change in either the atomic number or the mass number of the nucleus from which it is emitted. The nucleus just loses energy because of the emission.

The emission of all three types of radiation causes the resulting nucleus to be more stable as it possesses less energy than the original nucleus.

28.7 RADIOACTIVE SOURCES

The type of radiation emitted by a radioactive source depends on the nature of the source itself. The ${}^{90}\text{Sr}$ isotope emits β particles to become more stable whereas the ${}^{60}\text{Co}$ isotope becomes more stable by the emission of γ radiation. Although each radioactive isotope emits its own particular type of radiation it is possible to obtain sources which emit more than one type of radiation. In such sources the nucleus *produced* by the emission of one type of radiation is *itself* radioactive and emits its own characteristic radiation. In this way it is possible to obtain a number of different radioactive isotopes within the same source, each emitting its own characteristic radiation. Such a group of radioactive isotopes is shown below:



and a source initially consisting of ${}^{232}_{90}\text{Th}$ will eventually emit α , β and γ radiations, and contain all the isotopes shown above.

The strength of a radioactive source is measured in terms of its activity

The activity of any radioactive source equals the total number of nuclear disintegrations in the source in one second

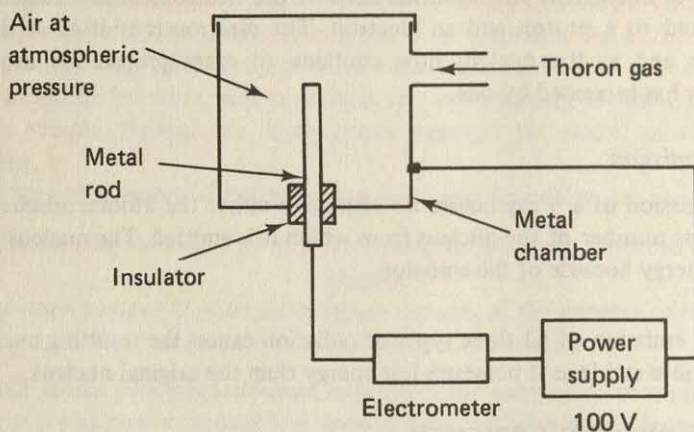
The activity is measured in curies and a source has an activity of one curie (1 Ci) if 3.7×10^{10} nuclear disintegrations occur within it in one second

28.8 RADIOACTIVE DECAY

The activity of radioactive sources decreases or *decays* with time.

The apparatus shown in Fig. 28.6 can be used to investigate the decay of the radioactive gas thoron. It consists of a metal chamber (called the

fig 28.6 *apparatus used to determine the half-life of radioactive thoron gas*



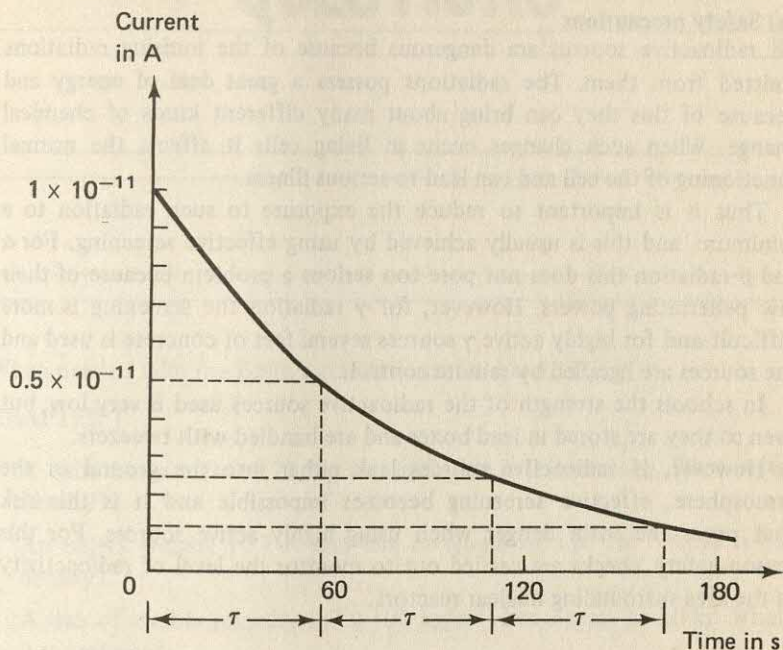
ionisation chamber), containing air at atmospheric pressure. A metal rod is placed through the base of the chamber and insulated from it. This rod is connected to one terminal of a power supply via an electrometer, with the other terminal of the supply connected to the chamber. The electrometer is able to detect very small currents of the order of 10^{-11} A. Initially no current is recorded in the circuit as the air in the chamber acts as an insulator.

However, when some radioactive thoron gas is introduced into the chamber a current is recorded on the electrometer. This current results from the α particles emitted from the thoron gas ionising the air in the chamber, and the size of the current depends on the number of ions produced. This in turn depends on the number of α particles emitted from the gas. If the activity of the gas is high, a large number of ions will be produced and a large current will be recorded on the electrometer. Thus, the size of the current in the circuit gives a measure of the activity of the gas and the variations in current show how the activity of the gas is changing.

Although the current is observed to fluctuate in a random way there is a steady drift to lower values. This indicates that although the α particles are emitted from the gas in a random way, there is a steady decrease in the number of particles that are being emitted every second. This decrease in the activity of the gas occurs because the number of radioactive nuclei in the chamber is steadily falling.

Fig. 28.7 shows how the current in the chamber varies with time. From the graph it can be seen that the current takes a set time τ to fall to half its original value no matter what original value is taken. Thus the decrease in the activity of the thoron gas occurs in a predictable way, halving its value

fig 28.7 the time τ for the activity to halve is always the same



in a fixed amount of time. It is found that this behaviour, i.e. halving in a fixed period of time, is common to all radioactive sources and the *half-life* is defined as follows:

The half-life of a radioactive source is the time it takes for the activity of the source to fall to half its original value (irrespective of what this value may be)

The value of the half-life depends on the source and it can have values ranging from millions of years to fractions of a second, and provided the half-life is known it is possible to predict how the activity of the source will decay.

As an example of this, consider the radioactive isotope of sodium $^{24}_{11}\text{Na}$ which has a half-life of 15 h. If a sample of this isotope is produced and has an activity of 500 microcuries (μCi), then after 15 h the activity will have dropped to around 250 μCi . Similarly, after a further period of 15 h the activity will be about 125 μCi . Although the *exact* value of the activity cannot be predicted, due to the random way in which the particles are emitted from the source, the half-life does enable an approximate value for the activity of the source to be predicted.

28.9 SAFETY PRECAUTIONS AND USES

(a) Safety precautions

All radioactive sources are dangerous because of the ionising radiations emitted from them. The radiations possess a great deal of energy and because of this they can bring about many different kinds of chemical change. When such changes occur in living cells it affects the normal functioning of the cell and can lead to serious illness.

Thus it is important to reduce the exposure to such radiation to a minimum, and this is usually achieved by using effective screening. For α and β radiation this does not pose too serious a problem because of their low penetrating powers. However, for γ radiation the screening is more difficult and for highly active γ sources several feet of concrete is used and the sources are handled by remote control.

In schools the strength of the radioactive sources used is very low, but even so they are stored in lead boxes and are handled with tweezers.

However, if radioactive sources leak either into the ground or the atmosphere, effective screening becomes impossible and it is this risk that poses the main danger when using highly active sources. For this reason safety checks are carried out to monitor the level of radioactivity in the area surrounding nuclear reactors.

(b) Uses

A radioactive isotope behaves chemically in an identical way to a non-radioactive atom of the same element, whilst still retaining its radioactive properties. This enables radioactive isotopes to be used as *tracers* to follow the path of a particular element through many physical or chemical processes. The use of small quantities of tracers mixed with a non-radioactive isotope of the element enables the assimilation of the element by a living organism to be followed, and the concentration of various chemicals in different organs can help in the diagnosis of various illnesses.

It is also possible to use radioactive sources in the treatment of various illnesses and this branch of medicine is called radiotherapy. It involves exposing the patient to small amounts of radiation, and such techniques are used in the treatment of cancer.

The absorption of radiation can be used to monitor the thickness of paper in paper mills and the thickness of steel in steel mills. The source is placed on one side of the paper and a Geiger-Müller tube on the other side. A change in the count rate indicates a change in the thickness.

QUESTIONS

When needed take the Earth's gravitational field as 10 N/kg

CHAPTER 1

- 1 A cube has a mass of 71.2 g and each side is 2.0 cm long. What is its density?
- 2 An object weighs 10.7 N and has a volume of 120 cm^3 . What is its density?
- 3 A slab of marble (density $2.7 \times 10^3 \text{ kg/m}^3$) has a mass of 10 kg . What is its volume?
- 4 A piece of plate glass has dimensions $30 \text{ cm} \times 40 \text{ cm}$ and a mass of 700 g . Taking the density of glass as 2.5 g/cm^3 , calculate the thickness of the glass.
- 5 A flask was weighed with a bung and clip and found to have a mass of 185.00 g . When the flask was evacuated the mass became 184.28 g . Using a measuring cylinder the flask volume was measured as 560 cm^3 . Calculate the density of air.
- 6 What is the mass of air in a room 6.0 m long, 4.5 m wide and 2.5 m high, given that the density of air is 1.29 kg/m^3 ?
- 7 Use the table of densities (Table 1.2) to answer this question.
 - (a) What is the radius of the base of a copper cylinder which is 1 cm long and has a mass of 90.0 g ?
 - (b) A spherical ball of ice has a radius of 2.0 cm . What is its mass?
 - (c) What is the side length of a cube of aluminium of mass 11.1 g ?
- 8 Copy and complete the table overleaf in order to answer the questions below.
 - (a) Which is the heaviest and which the lightest object?
 - (b) Which object has the smallest and which the largest volume?
 - (c) Which object is made of the most dense material?

<i>Object</i>	<i>Mass</i>	<i>Volume</i>	<i>Density</i>
A	17.8 kg	2.0 m ³	
B		2.0 m ³	920 kg/m ³
C	6.0 g		2.7 g/cm ³
D	1.0 g		0.24 g/cm ³
E		350 cm ³	18 700 kg/m ³

CHAPTER 2

- 1
 - (a) What is the pressure exerted when a force of 10 N acts over an area of 2 m²?
 - (b) What pressure is caused by a force of 30 N acting on an area of 5 m²?
 - (c) The pressure at a certain depth in a liquid is 1000 Pa; what total force does it exert on the base of its container, of area 2 m², which is all at this depth?
 - (d) Taking atmospheric pressure as 1.0×10^5 Pa, calculate the force exerted on one face of a rectangular plate glass window of dimensions 3 m $\times$ 2 m.
- 2 A rectangular block has dimensions 4.0 m $\times$ 2.0 m $\times$ 1.0 m. It weighs 50 kN.
 - (a) What is the greatest pressure it can exert on a horizontal bench?
 - (b) What is the least pressure it can exert while resting on the same bench?
- 3 A model hovercraft has a mass of 30 kg and the dimensions of its base are approximately 20 cm by 40 cm. What is the average pressure which the hovercraft exerts on the ground?
- 4 Pressure is defined as force/area. A point has no area. What do you think we mean when we talk about pressure at a point?
- 5 Give an explanation for each of the following:
 - (a) A knife with a very sharp blade will cut very much more easily than a knife with a very blunt blade.
 - (b) A drawing pin can be pushed into some wood fairly easily, whereas a metal bar cannot be pushed into the wood.
 - (c) Toboggans with very thin runners leave more impression in the snow than toboggans with wide runners.
 - (d) Tractors have tyres of very large area in contact with the ground.
- 6
 - (a) A girl weighing 500 N is wearing shoes which have very small heels (stiletto heels). The area of each heel is 0.5 cm² and the

area of each sole in contact with the ground is 50 cm^2 . What is the pressure she is exerting on the floor if

- (i) she stands with her weight equally distributed between both feet,
 - (ii) she stands on one foot,
 - (iii) she balances on one heel only?
- (b) An elephant weighs 60 kN and each of its feet has an area of 0.07 m^2 in contact with the ground. What is the pressure the elephant exerts on the ground, if
- (i) it stands on all four feet,
 - (ii) it balances on two feet?
- 7 The Evesham pressure apparatus consists of a polythene bag inside a wooden box that has two loosely fitting lids. Lid Y is four times the area of lid X.
- (a) A 10 N weight is placed on lid X and two 10 N weights are placed on lid Y. Which lid rises first when the polythene bag is inflated?
 - (b) What happens when the bag is inflated if lid X has one 10 N weight on it and lid Y has four 10 N weights on it?
 - (c) Suppose the pressure inside the polythene bag is 100 Pa , what is the force on 0.5 m^2 of surface of the bag?
 - (d) Suppose that lid X has an area of 0.02 m^2 and lid Y an area of four times this. If the pressure in the bag remains at 100 Pa , what is the force upwards on lid X? What is the force upwards on lid Y?
- 8 Suppose that you see an accident in which someone skating on the ice on a frozen lake suddenly falls through into the water because the ice is too thin and breaks. Two possible actions occur to you. One is to run for a ladder which you see some distance away and use this as a means of setting out across the ice. The other is to run out straight away on the ice to get to the person as soon as possible. Which method is more likely to result in a successful rescue, and why?
- 9 Consider a tank with a square cross-section of area $A \text{ m}^2$. Suppose it is filled to a depth $h \text{ m}$ with a liquid of density $d \text{ kg/m}^3$.
- (a) What is the volume of liquid which fills the tank?
 - (b) What is the mass of this liquid which fills the tank?
 - (c) What is the weight of this liquid which fills the tank? (Remember that a mass of 1 kg has a weight of 10 N .)
 - (d) This weight acts over the base of the tank. Over what area does it act?
 - (e) What is the pressure acting on the base?

These steps should have enabled you to prove the following

relationship:

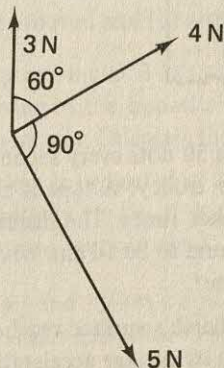
$$\text{Pressure (Pa)} = 10 \text{ (N/kg)} \times \text{Depth, } h \text{ (m)} \times \text{Density, } d \text{ (kg/m}^3\text{)}$$

- 10 Give an explanation for both of the following:
 - (a) In order to get liquid out of a sealed tin one must either pierce two small holes on either side of the tin or make one very large hole.
 - (b) The pressure of the atmosphere decreases as the altitude increases.
- 11 Suppose you are commissioned to write a textbook for the inhabitants of a planet where the atmospheric pressure is 100 mmHg (on Earth it is about 760 mmHg). How would the chapter you wrote on atmospheric pressure differ from those written for use on Earth?
- 12 A bottle of lime juice is corked at sea level when the atmospheric pressure is 800 mmHg. It is then taken up a mountain of height 1080 m. If the atmospheric pressure decreases by 10 mmHg for a rise of 120 m, what is the pressure tending to push the cork out of the bottle at the top of the mountain? If the diameter of the cork is 2 cm, what is the resultant force on the cork? (Density of mercury = 13600 kg/m^3 .)
- 13 The following questions refer to an aneroid barometer.
 - (a) Why is the box corrugated?
 - (b) Why is the complicated lever system necessary?
 - (c) What is the purpose of the hair spring?
- 14
 - (a) Describe, giving full experimental details, how a simple mercury barometer could be set up.
 - (b) Does a simple mercury barometer demonstrate the fact that the Earth's atmosphere exerts a pressure? Give reasons for your answer.
 - (c) Describe some other experiment to illustrate that the Earth's atmosphere exerts a pressure.
 - (d) What effect, if any, would be produced in the vertical height of the mercury in a simple mercury barometer if
 - (i) the bore of the tube were changed,
 - (ii) the tube were tilted to an angle of 30° with the vertical,
 - (iii) some air were introduced in the space above the mercury?
- 15 A hydraulic machine is required to lift a load of 1000 N. The pressure available to operate the machine is 250 Pa. What must be the area of the piston on which the load is standing?

CHAPTER 3

- 1 Determine the resultant (in magnitude and direction) of the following pairs of forces:
 - (a) 3 N and 4 N acting at right-angles to each other,
 - (b) 10 N and 20 N acting at 60° to each other,
 - (c) 15 N and 30 N acting at 45° to each other.
- 2 An aeroplane flies due south at 200 km/h and a wind is blowing from east to west at 40 km/h. What is the resultant velocity of the aeroplane?
- 3 A man of weight 800 N is water-skiing. The rope pulling him is horizontal and the tension in it is 200 N. What is the resultant of these two forces acting on him? In what direction does it act?
- 4 A string has a mass of 100 g attached to its end and is hanging vertically. What horizontal force must be applied to the mass to produce an inclination in the string of 30° to the vertical? What is the tension in the string in this case?
- 5 Two tugs are manoeuvring a supertanker by hawsers attached to its bow. Each tug pulls with a force of 40 000 N and the angle between the hawsers is 30° . What is the resultant force on the supertanker?
- 6 Explain why a slack clothes line is less likely to break than a taut one.
- 7 A boat is rowed across a stream. The boat is always pointing in a direction which is perpendicular to the direction of flow of the stream. The speed of the boat would be 5 km/h in still water. If the stream is flowing at 3 km/h, what is the speed and the direction of travel of the boat?
- 8 Forces of magnitudes 3 N, 4 N and 5 N act at a point. The angle between the 3 N force and the 4 N force is 60° , and the angle between the 4 N and the 5 N force is 90° (Fig. 3.8). What is the resultant force?

fig 3.8

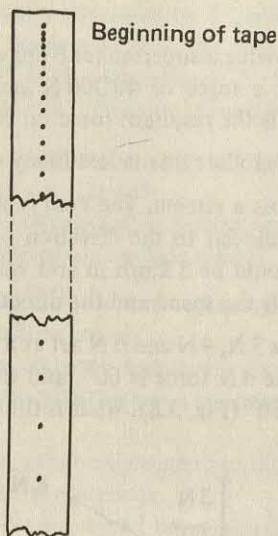


- 9 A truck is pulled along some rails by a rope. When the rope is horizontal and parallel to the rails, the force required is 100 N. What force would be required to pull it along at the same velocity if the rope made an angle of 30° with the rails?

CHAPTER 4

- 1 Describe what experiment you would carry out to determine how many dots were made in a second by a ticker timer.
- 2 Describe the motion of the body represented by the dots in Fig. 4.18. If the ticker timer makes 50 dots every second, what is the velocity at the beginning of the motion and what is the velocity at the end of the motion?

fig 4.18



- 3 A ticker timer produces 50 dots every second and is used to determine the average speed of a trolley. A tape is attached to the trolley and passes through the ticker timer. The distance between the third and thirteenth dots was found to be 10 cm. What was the average speed of the trolley over this time?
- 4 A rocket leaving the Earth's surface reaches a speed of 36 000 km/h in two minutes. What is its average acceleration?

- 5 A spacing of successive dots on a tape which has passed through a ticker time are 0.7, 0.9, 1.1 and 1.3 cm. The ticker timer is making 50 dots every second. What is the acceleration of the trolley?
- 6 A force of 10 N is applied to a mass of 2 kg. What acceleration is produced?
- 7 Two men are pushing a car of mass 1000 kg along a level road at a constant speed. When a third man joins them the car accelerates at 0.1 m/s^2 . If all the men push with the same force, with what force does each man push?
- 8 Sketch ticker timer tapes, showing 10 consecutive dots, which you would obtain from the following experiments:
 - (a) a trolley accelerating at 4 cm/s^2 with a ticker timer that makes 30 vibrations per second,
 - (b) a key-ring falling vertically which is attached to a tape passing through a ticker timer making 50 dots every second,
 - (c) a trolley of mass 2 kg which is accelerated by a force of 2 N.
- 9 A boy pulls a cart of mass 100 kg along a horizontal road. He pulls with a force of 250 N. If the cart starts from rest and neglecting friction, find
 - (a) the acceleration of the cart,
 - (b) how far the cart moves in the first second,
 - (c) how far the cart would move in 20 s.Would it be possible for the boy to exert this force for 20 s?
- 10 A man of mass 100 kg lives at a place where the gravitational constant is 9.80 N/kg . He moves to another place in the world where the Earth's gravitational constant is 9.77 N/kg .
 - (a) What is his change in weight due to the move?
 - (b) How fast would he accelerate if he jumped off the roof of a shed (a) before he moved and (b) after he moved?
- 11 A trolley X of mass 2 kg moving at 3 m/s collides with a trolley Y of mass 1.0 kg initially moving in the opposite direction at 1.0 m/s. The trolleys stick together on impact. Calculate their common final velocity.
- 12 Trolley Y (of mass 1.0 kg) is projected at 2.0 m/s towards trolley X (mass 2.0 kg) which is initially at rest. If the trolleys stick together on impact, calculate their subsequent velocity.
- 13 A rifle has a mass of 5 kg and it fires a bullet of mass 25 g with a velocity of 100 m/s. What is the initial recoil speed of the rifle when the bullet is fired?

- 14 This question is about the equations in Section 4.12. Use the definition of acceleration to show that for a body starting with velocity u

$$v = u + at$$

Now substitute in the equation,

$$\text{Average velocity} = (\text{Initial velocity} + \text{Final velocity})/2$$

and show that

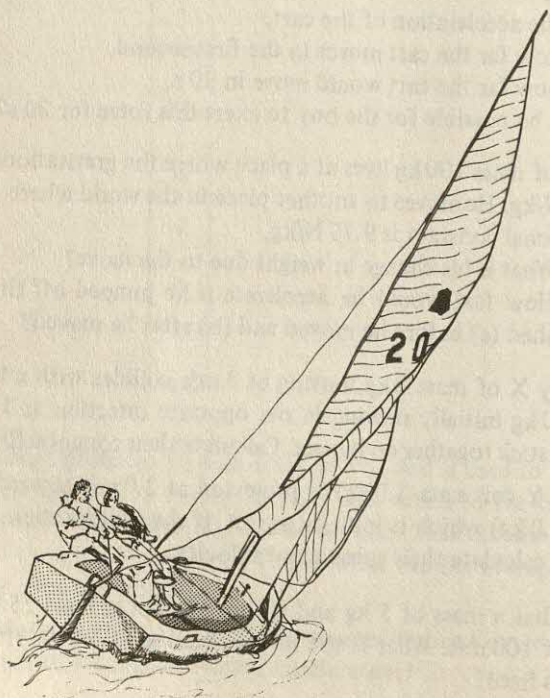
$$s = ut + \frac{1}{2}at^2$$

Finally by eliminating t from $v = u + at$ and $s = ut + \frac{1}{2}at^2$, show that

$$v^2 = u^2 + 2as$$

A vehicle is moving at 2 m/s. It accelerates at 3 m/s², what is its velocity after it has travelled 6 m?

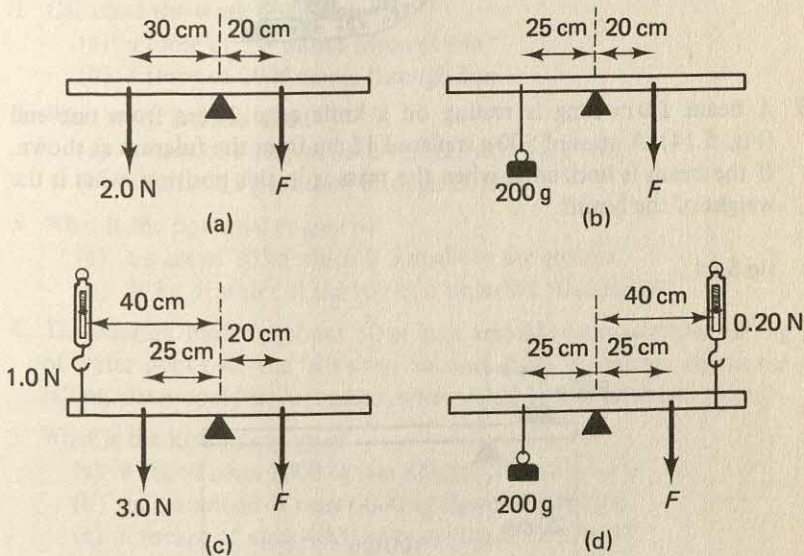
fig 5.11



CHAPTER 5

- 1 At one of the sideshows at a fair, Betty and Jim are throwing balls at skittles on a shelf. Jim says that there is more likelihood of knocking a skittle off the shelf if you hit it in the middle, whereas Betty says he's wrong and that there is more likelihood of knocking it off the shelf if you hit it near the top. Which of them is right, and why?
- 2 Why is the sailing dinghy (Fig. 5.11) less likely to capsize if the occupants are leaning over the side as shown?
- 3 Each of the uniform beams shown in Fig. 5.12 is balanced with the fulcrum at the centre of the beam. In each case calculate the magnitude of the unknown force F if the beam is to remain horizontal.

fig 5.12

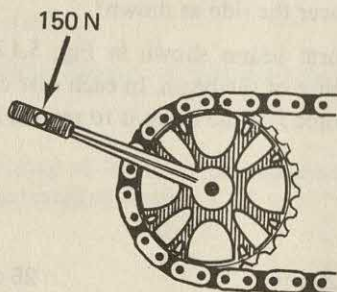


- 4 A girl who weighs 300 N sits 2.0 m from the fulcrum of a seesaw. A boy weighing 450 N is seated on the other side. How far from the fulcrum is the boy if the seesaw remains horizontal?
- 5 Jim weighs 600 N and sits 2.0 m from the centre of a seesaw. Mary weighs 500 N . Where must she sit if the seesaw is to remain horizontal? If Mary moves 1.4 m nearer to the fulcrum, where must Helen, weighing 300 N , sit in order to keep the seesaw horizontal?

- 6 A man riding a bicycle applies a force of 150 N to the pedal. If the force is applied vertically downwards, what is the moment of the force about the axis of the crank (a) when the crank is vertical and (b) when the crank is horizontal? (Take the crank length as 16 cm.)

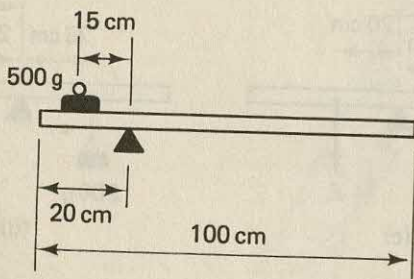
Would it be sensible for a cyclist to try always to turn the pedals so that he pushes at right-angles to the crank (Fig. 5.13)? Explain.

fig 5.13



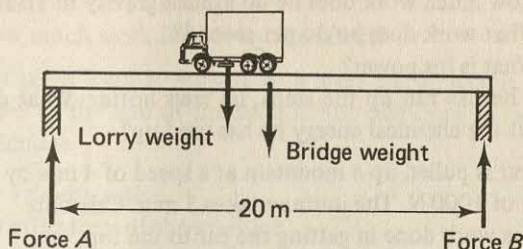
- 7 A beam 1.0 m long is resting on a knife edge 20 cm from one end (Fig. 5.14). A mass of 500 g is placed 15 cm from the fulcrum as shown. If the beam is horizontal when the mass is in this position, what is the weight of the beam?

fig 5.14



- 8 A uniform bridge 20 m long has a mass of 30×10^3 kg and is supported at each end (Fig. 5.15). A lorry which has a mass of 3.0×10^3 kg is on the bridge. How much weight is carried by each of the supports (a) when the lorry is on the centre of the bridge and (b) when the lorry is 4 m from one end? (Hint: one way to solve this problem is to take moments about the point where one of the supports makes contact with the bridge.)

fig 5.15 *the forces A and B must act with the bridge weight and lorry weight so as to keep the bridge from moving or turning*



CHAPTER 6

(Take the force of gravity as 10 N/kg and the acceleration due to gravity as 10 m/s^2 .)

- 1 Calculate the work done when
 - (a) a force of 3 N moves through 4 m ,
 - (b) a force of 20 N moves through 5 m .
- 2 What work is done when
 - (a) a mass of 10 kg is raised vertically through 2 m ,
 - (b) a mass of 5 g is raised through 20 m ?
- 3 What is the potential energy of
 - (a) a mass of 20 kg which is 5 m above the ground,
 - (b) 20 kg of water at the top of a waterfall 50 m high?
- 4 The Niagara Falls are about 50 m high and it is estimated that 10^7 kg of water pour over the falls every second. If all the energy due to the falling water could be harnessed, what would be the available power?
- 5 What is the kinetic energy of
 - (a) a car of mass 1000 kg travelling at 30 km/h ,
 - (b) an aeroplane of mass 6000 kg flying at 300 m/s ,
 - (c) a rocket of mass 4000 kg travelling at 2000 m/s ?
- 6 A man pushes a frictionless trolley of mass 50 kg across a room, the trolley being initially at rest. After it has been pushed 3 m it is moving at 2 m/s . With what average force has he pushed the trolley?
- 7 State the principle of conservation of energy and explain how it applies to
 - (a) a pendulum swinging,
 - (b) a squash ball being hit by a racket,
 - (c) a space shuttle re-entering the Earth's atmosphere and eventually landing on an airport runway.

- 8 A boy runs up steps through a height of 20 m in 10 s. His weight is 600 N.
- How much work does he do against gravity in 10 s?
 - What work does he do per second?
 - What is his power?

After he has run up the steps, he feels hotter. What does this tell you about the chemical energy he has used up?

- 9 A cable car is pulled up a mountain at a speed of 4 m/s by a cable with a tension of 1000 N. The journey takes 5 min. Calculate
- the work done in getting the car to the top,
 - the work done in getting the car to the top if the speed of travel had been only 2 m/s.

What power was needed for the engine to achieve the higher speed of climb?

- 10 Two boys are discussing roller-coaster rides. One says that each summit on the track must be lower than the previous one. The other boy says that as long as the first summit is highest then it does not matter what height any of the other ones have.

Discuss their statements.

CHAPTER 7

(Where necessary take the Earth's gravitational field as 10 N/kg.)

- An electric motor raises a 10 kg mass through a distance of 2 m. In doing this it consumes 250 J of energy. What is the efficiency of the motor?
- A pulley system has two pulleys on the top block and two pulleys on the bottom block. A load of 1000 N is hung from the bottom block. Draw the pulley system.

It is found that an effort of 300 N is needed to raise the load.

- How much energy is supplied to the machine by the effort when the effort moves through 5 m?
- How far does the load rise when the effort moves through 5 m?
- How much energy is gained by the load when the effort moves through 5 m?
- What is the numerical ratio

$$\frac{\text{Energy transferred to the load}}{\text{Energy supplied to the machine by the effort}}?$$

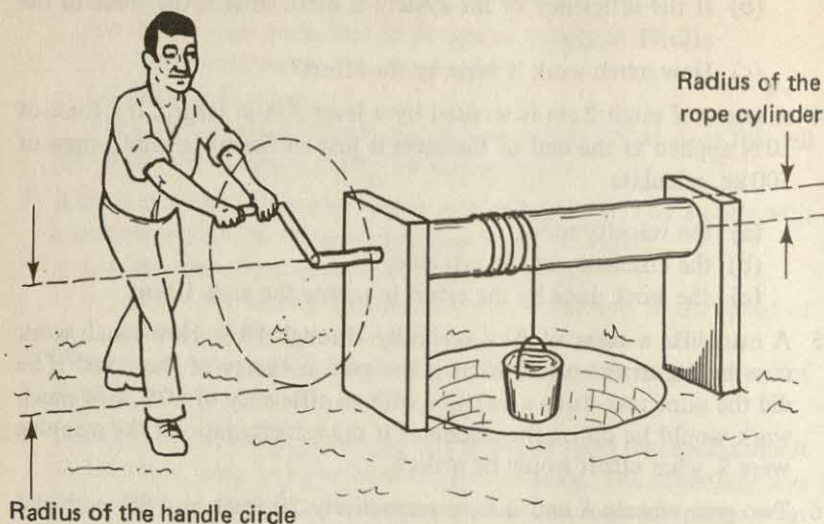
- What is the ratio called?
 - Could this ratio be greater than 1? Explain your answer.
- 3 An inclined plane is used to raise a load of 500 N through a height of 3 m. The load moves 12 m along the plane.

- (a) How much work is done on the load?
 - (b) If the efficiency of the system is 80%, what is the value of the effort?
 - (c) How much work is done by the effort?
- 4 A screw of pitch 2 cm is worked by a lever 3 m in length. If a force of 10 N applied at the end of the lever is just sufficient to raise a mass of 500 kg, calculate
- (a) the velocity ratio,
 - (b) the efficiency of the machine,
 - (c) the work done by the effort in raising the mass 10 cm.
- 5 A man lifts a mass of 6 kg vertically through 10 m. How much work does he do on the mass? What is the gain in energy of the mass? If he did the same task using a machine with an efficiency of 60%, how much work would he do on the machine? If the velocity ratio of the machine were 5, what effort would he make?
- 6 Two gear wheels A and B have respectively 20 teeth and 80 teeth and are meshed together. If A is driven by a motor and itself drives B, what is the velocity ratio of the arrangement? Where might you find gear wheels used in this way? If the arrangement were that wheel B was driving wheel A, where might this arrangement be found?
- 7 Write down equation (7.1). For 'Work got out of machine' substitute 'Load $\times$ Distance load moves'. For 'Work put into machine' substitute 'Effort $\times$ Distance effort moves'. Hence show that

$$\text{Efficiency} = \frac{MA}{VR}$$

- 8 A windlass (Fig. 7.8) is used to lift a bucket from a well. The cylinder on which the rope winds has a radius of 20 cm and the handle has an effective radius of 50 cm. The bucket has a mass of 2 kg and holds 18 kg of water which it raises through 10 m. The efficiency of the system for raising water is 50%.
- (a) What is the velocity ratio of the system?
 - (b) What is the least force which must be applied to the handle in order to raise the load?
 - (c) What is the work done by the effort in raising a bucket full of water?
 - (d) One way of decreasing the force needed to turn the handle is to attach a second bucket to the rope cylinder. Draw a diagram showing clearly how this would be done and calculate the effort needed using this arrangement if both the buckets have the same mass.

fig 7.8 the windlass



CHAPTER 8

(Where necessary take the Earth's gravitational field as 10 N/kg and the density of water as 1000 kg/m^3 .)

1 Explain the following statements:

- If a water-skier is being hauled out of the water into the ski boat, the skier appears to get heavier as he is lifted out of the water.
- If you are swimming in the sea on a shingle beach, the stones hurt much less as you run into the water.
- If an egg is placed in a saucepan full of water, it sinks to the bottom; but when a lot of salt is added to the water and the water is stirred until the salt dissolves, the egg floats to the surface.
- When a ship sailing from a river into the sea enters the sea, the level of the water around its hull changes. Does the water level rise or fall?
- When you step from the shore into a canoe, the canoe sinks lower in the water.
- Iron is denser than water and sinks if placed in water. Yet an iron ship floats in water.
- A balloon filled with hydrogen rises but a balloon filled with air falls.

2 An object of density 8 g/cm^3 has a volume of 10 cm^3 .

- What is the mass of the object?
- What is its weight in the air?

- (c) What is its apparent weight when completely immersed in water?
- (d) What is its apparent weight when completely immersed in paraffin of density 0.8 g/cm^3 ?
- 3 This question refers to the data given in Table 8.1.
- What is the weight of water displaced by the stone (see table)?
 - What is the mass of water displaced by the stone?
 - What is the volume of water displaced by the stone?
 - What is the volume of the stone?
 - What is the volume of paraffin displaced by the stone?
 - What is the weight of paraffin displaced by the stone (see table)?
 - What is the mass of paraffin displaced?
 - From the answers to parts (e) and (g) calculate the density of paraffin.
 - By an exactly similar process calculate the density of salt solution.
- 4 A cube of wood of sides 10 cm has a mass of 0.6 kg . What is
- the density of the wood,
 - the weight of the wood in air,
 - the weight of water displaced by the wood when it is floating,
 - the volume of water displaced by the wood when it is floating,
 - the force which must be applied to the wood in order that it becomes just completely immersed?
- 5 A balloon of volume 1.0 m^3 is filled with hydrogen. What is the up-thrust and the resultant force on the balloon? (You may neglect the weight of the fabric. The density of air may be taken as 1.3 kg/m^3 and the density of hydrogen as 0.09 kg/m^3 .)
- 6 A character called Pyecraft in one of H. G. Wells's science fiction novels has a body which loses all its weight but which retains its original volume. In the novel he floats fully clothed to the ceiling. Is this a plausible consequence of Archimedes' principle? Give reasons for your answer.
- 7 A family is taking a boat out on a lake. They have a picnic basket which is full of tins of food. If the picnic basket drops over the side of the boat, will the level of the water in the lake rise or fall? Give reasons for your answer.
- 8 A polar bear of mass 800 kg is standing on an iceberg. The density of ice is 920 kg/m^3 . The iceberg drifts into warmer water and begins to melt. If the density of the water is 1200 kg/m^3 , what is the least volume of ice remaining before the polar bear gets his feet washed by the sea?
- 9 A flat bottomed test-tube of uniform cross-section 0.5 cm^2 is weighted by putting some lead shot in it so that it floats vertically in water. How far will it sink in water if the total mass of the test-tube and contents is 2.5 g ? How far will it sink in a liquid of density 0.8 g/cm^3 ?

CHAPTER 9

(Where needed, take the specific heat capacity of water as 4200 J/kg K .)

- 1 (a) What is the specific heat capacity of paraffin if the temperature of 0.2 kg of it rises by 10 K when an electric heater supplies it with 4400 J of heat energy?
 (b) A 3 kW heater is used to heat 30 kg of water. What is the maximum possible temperature rise in 3 min ?
- 2 In an experiment using the apparatus illustrated in Fig. 9.1, a force of 30 N moved through 3 m and the copper cylinder of mass 100 g rose in temperature from 15°C to 18°C . What is the specific heat capacity of copper?
- 3 A metal block of mass 5 kg is dragged over a surface by a force of 10 N . If it is pulled a distance of 100 m , how much work is done against friction? If half this energy goes into the block, by how much will its temperature rise if its specific heat capacity is 400 J/kg K ?
- 4 A waterfall is 600 m high. How much kinetic energy is gained by 1 kg of water as it falls? By how much will the temperature of water at the bottom of the fall exceed that at the top?
- 5 The hot water system in a house is not functioning properly, and when the water is run into a bath, the resulting temperature of the bath is only 30°C . If the bath contains 0.12 m^3 of water at this temperature, how many saucepans, each containing $2.0 \times 10^{-3} \text{ m}^3$ of boiling water, will need to be added to the bath, in order to raise its temperature by 5 K ? Neglect heat losses and give your answer to the nearest whole number.
- 6 A copper calorimeter of mass 80 g contained 120 g of oil at 15°C . A piece of copper of mass 30 g was transferred quickly from a furnace to the calorimeter. The temperature of the oil rose to 45°C . Find the temperature of the furnace. State any possible sources of error. (Specific heat capacity of copper = 380 and that of the oil = 1900 , both in J/kg K .)
- 7 A hydro-electric power station is fed from a reservoir of area $1000 \text{ m} \times 1000 \text{ m}$ and 50 m deep. The reservoir has its bottom 300 m above the power station generators. At peak load the power station must supply 200 MW .
 - (a) What is the total potential energy of the water in the reservoir?
 - (b) At what rate is the water flowing through the generating turbines when the station is operating at its peak load?
 - (c) What is the maximum time for which the station can produce electrical energy at 200 MW ?

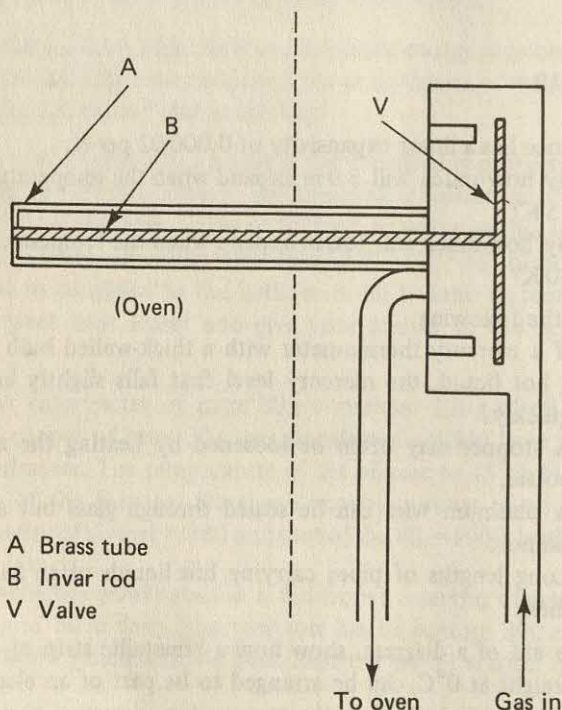
- (d) In practice the rate of water flow needed for peak power output must be greater or less than that calculated in (b). Which must it be, and why? (Density of water = 1000 kg/m^3).
- 8 In a solar furnace a mirror is used to focus the Sun's rays onto a metal block. The temperature of the block rises 10°C in 5 min. The mass of the block is 10 kg and the specific heat of the material of which it is made is 5000 J/kg K .
- How much energy is received by the block every minute?
 - How much heat is received by the block in 10 h, assuming a constant rate of supply?
 - Supposing that the area of the mirror collecting the Sun's rays was 1.0 m^2 , how much energy could be collected from an area of desert $100 \text{ m} \times 100 \text{ m}$ in 10 hours?
 - If this energy were used to drive electric generators and the system were 50% efficient, what average power could be produced over the 10 h period?

CHAPTER 10

- 1 A substance has a linear expansivity of 0.00002 per K.
- By how much will 5.0 m expand when the temperature is raised 15 K ?
 - By how much will 100 m expand when the temperature is raised 30 K ?
- 2 Explain the following:
- If a mercury thermometer with a thick-walled bulb is put into a hot liquid, the mercury level first falls slightly before rising quickly.
 - A stopper may often be loosened by heating the neck of the bottle.
 - A platinum wire can be sealed through glass but a brass one cannot.
 - Long lengths of pipes carrying hot liquids often have loops in them.
- 3 With the aid of a diagram, show how a bimetallic strip of copper and invar, straight at 0°C , can be arranged to be part of an electric circuit to ring an electric bell when the temperature rises to a particular value above 0°C .
- How could the system be adjusted to make the bell ring at
- a higher temperature,
 - a temperature below 0°C ?

- 4 A bar is 100 cm long at 10°C and expands 0.0018 cm when heated to 100°C . What is the linear expansivity of the material?
- 5 Explain carefully under what circumstances the temperature of water in a pond may
 - (a) be higher at the surface than at the bottom,
 - (b) be higher at the bottom than at the surface.
- 6 A mercury thermometer being checked is found to read 101.0°C at the upper fixed point and -0.5°C at the lower fixed point. What is the correct temperature when the thermometer reads 20.0°C ?
- 7 Fig. 10.9 is a diagram of a gas thermostat as fitted, for example, on a gas oven. Remembering that brass expands much more than invar if both are equally heated, explain carefully how the device works to keep the oven temperature approximately constant.

fig 10.9 a simplified diagram of a thermostat for a gas oven



- 8 Give *two* examples where the expansion of a solid when heated is made use of and *two* where expansion is a difficulty which has to be allowed for. In the latter cases, explain exactly how the allowance is made.

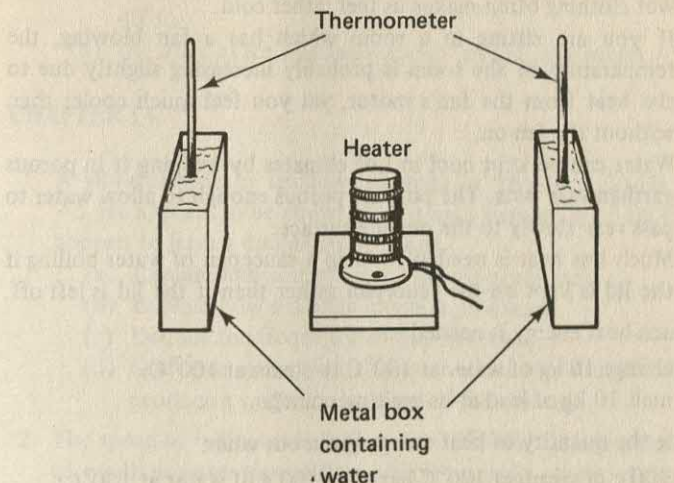
CHAPTER 11

- 1 What is the distinction between the three states of matter on the molecular theory?
- 2 What is Brownian motion? Describe, with the aid of a diagram or diagrams, the apparatus you would set up to demonstrate Brownian motion of either smoke particles suspended in air or particles suspended in a liquid.
- 3 What evidence is there to suggest that the molecules of a liquid are closer together than the molecules in the corresponding gas?
- 4
 - (a) Use the kinetic theory to explain why a gas exerts a pressure on the walls of its container.
 - (b) Show how the kinetic theory leads to an explanation of Boyle's law.
- 5 Explain, in terms of the kinetic theory of gases:
 - (a) the change in pressure of a gas which is compressed at constant temperature,
 - (b) the change in pressure when a gas is heated at constant volume,
 - (c) the increase in temperature when a fixed amount of gas, say in a bicycle pump, is compressed and not allowed to escape.
- 6 At a pressure of 10^5 N/m^2 the volume of a certain mass of gas is 100 cm^3 . What will be the pressure if the volume increases to 600 cm^3 , the gas temperature remaining constant?
- 7 At each stroke of a pump, 500 cm^3 of air at atmospheric pressure are compressed and delivered to a car tyre. The initial pressure in the tyre is 2.0 atmospheres. If the volume of the tyre remains constant at 20 litres and assuming that the temperature remains constant, calculate the new pressure after 10 strokes of the pump.
- 8 The pressure of air in a car tyre is $1.5 \times 10^5 \text{ N/m}^2$ when the temperature is 20°C . What will the pressure become if the temperature rises to 25°C , assuming that the tyre volume remains unchanged?
- 9 At 27.0°C and 2.0 atmospheres pressure the density of a gas is 0.40 g/dm^3 . What will its density be at 4.0 atmospheres pressure and 227.0°C ?
- 10 A vessel used for storing gas has a safety valve which blows off at 10^6 N/m^2 . It contains gas at a pressure of $8.0 \times 10^5 \text{ N/m}^2$ at 15°C . At what temperature would the valve start to blow off?

CHAPTER 12

- 1 Explain the following:
 - (a) A vacuum is the worst possible conductor of heat.
 - (b) Double glazing keeps rooms warm in winter and cool in summer.
 - (c) In hot weather, ice cream will take a long time to melt if newspaper is wrapped round it.
 - (d) There is frequently a cold draught below a closed window.
 - (e) Pouring boiling water into a tumbler made of thick glass is liable to crack the glass, but if boiling water is poured into a tumbler of thin glass the glass is unlikely to crack.
 - (f) Pipes carrying cold water in the home in winter time are lagged.
 - (g) Eskimos can keep themselves warm in a cold climate by building huts made of ice.
 - (h) A fan rotating in a room may cause you to feel much cooler but a thermometer in the room will show the same temperature as before the fan was switched on.
 - (i) The ice compartment in a refrigerator is placed at the top of the refrigerator compartment.
 - (j) Fur coats would not look as nice if worn inside out but they would be much more effective in keeping the occupant warm.
 - (k) Heat energy from the Sun has been falling on the Earth for many millions of years, yet the temperature of the Earth remains approximately constant.
- 2 Describe on a molecular and atomic scale how heat energy is conducted along a metal rod.
- 3 Fig. 12.11 is a diagram showing the principles of a hot water system in a house. Discuss why each pipe which joins the hot water tank joins at the top or bottom of the tank.
- 4 Explain why a coal or log fire which uses a chimney helps to ventilate a room.
- 5 What part do conduction, convection and radiation play in cooling a motor car engine?
- 6 How does a pie being cooked in a heated oven receive its heat? Discuss this in terms of conduction, convection and radiation and explain by which method most of the heat reaches the pie. How does the centre of the pie become warm?
- 7 Fig. 12.12 illustrates an experiment in which an electric heater is placed midway between two metal containers which have water and a thermometer in them. What would you notice about the readings on the thermometers when the heater is switched on if

fig 12.12 *the metal containers are identical in size and each contains the same amount of water*



- (a) both the metal containers were made of shiny copper,
 - (b) the one on the left were made of shiny copper and the one on the right had been blackened by holding it in a candle flame,
 - (c) the one on the left were painted with a white gloss paint and the one on the right were painted with a matt black paint?
- 8 What explanation can you give for the fact that you can sit indoors on a sunny day and feel the warmth of the Sun, yet if you put a glass fire-screen in front of a fire very little heat reaches you.
- 9 Describe the construction of a Thermo's flask.
Explain how heat losses by conduction, convection and radiation are eliminated or reduced. How in fact is any heat lost from the flask? Is the flask just as effective for storing cold liquids as hot liquids?

CHAPTER 13

(Where necessary use the following constants: specific heat capacity of water = 4200 J/kg K ; specific latent heats of fusion of ice and lead are $3.3 \times 10^5 \text{ J/kg}$ and $2.1 \times 10^4 \text{ J/kg}$ respectively; specific latent heat of vaporisation of water = $2.3 \times 10^6 \text{ J/kg}$.)

- 1 Discuss each of the following statements:
- (a) A scald from steam is very much worse than one from boiling water.

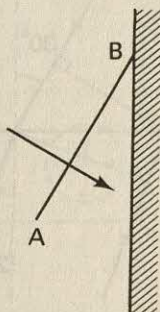
- (b) Skating on ice becomes very difficult if the temperature is very low.
 - (c) Wet clothing often makes us feel rather cold.
 - (d) If you are sitting in a room which has a fan blowing, the temperature of the room is probably increasing slightly due to the heat from the fan's motor, yet you feel much cooler than without the fan on.
 - (e) Water can be kept cool in hot climates by keeping it in porous earthenware pots. The pots are porous enough to allow water to pass *very* slowly to the outside surface.
 - (f) Much less heat is needed to keep a saucepan of water boiling if the lid is kept on the saucepan rather than if the lid is left off.
- 2 How much heat energy is needed to
- (a) change 10 kg of water at 100°C to steam at 100°C ;
 - (b) melt 10 kg of lead at its melting point?
- 3 Calculate the quantity of heat energy given out when
- (a) 100 g of steam at 100°C becomes 100 g of water at 100°C ;
 - (b) 100 g of water at 100°C becomes 100 g of water at 0°C ;
 - (c) 100 g of water at 0°C becomes 100 g of ice at 0°C ;
 - (d) 100 g of steam at 100°C becomes 100 g of ice at 0°C .
- 4 How long will it take a 3 kW immersion heater completely to melt 10 kg of ice at 0°C ?
- 5 How long will a kettle rated at 3 kW take to bring 1.0 kg of water at 10°C to the boil? If the kettle is left on after the water reaches its boiling point, how much water will boil away in 3 min?
- 6 A 1 kW kettle contains 0.5 kg of water at 20°C . It is switched on for 12 min and after this time 0.3 kg of water are left in the kettle. How much heat energy has been wasted, that is, not supplied to the water?
- 7 200 g of ice are placed in a beaker. A bunsen burner is placed under the beaker and all the ice has melted in six minutes.
- (a) Calculate the heat energy supplied to the beaker per minute.
 - (b) How long after all the ice has melted would it take for the water to reach its boiling point?
 - (c) What time would elapse after the water began to boil before it had all boiled away?
- 8 A calorimeter of mass 100 g and specific heat capacity 400 J/kg K contains some ice floating in 150 g of water. After 15 g of steam is condensed in the water, the final temperature is 40°C . Calculate
- (a) the heat energy that is needed to raise the temperature of the calorimeter by 1°C ,

- (b) the heat needed to turn 15 g of steam at 100°C to water at 100°C ,
- (c) the heat needed to turn 15 g of steam at 100°C to water at 40°C ,
- (d) the mass of ice originally present in the calorimeter.

CHAPTER 14

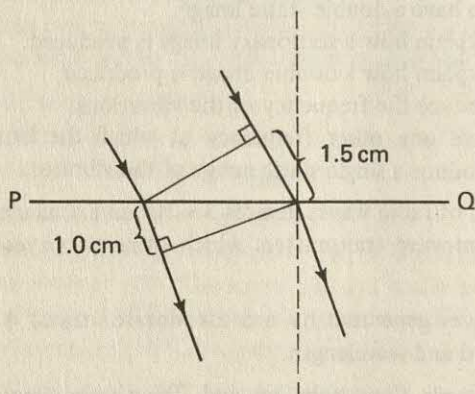
- 1 A vibrator seen in stroboscopic light of frequency 30 Hz, 15 Hz, or 7.5 Hz appears to be at rest: but in light flashing at 60 Hz the vibrator appears to have a double static image.
 - (a) Explain how a stationary image is produced.
 - (b) Explain how a double image is produced.
 - (c) Deduce the frequency of the vibrations.
 - (d) Give one other frequency at which the lamp flashes would produce a single static image of the vibrator.
- 2 The speed of radio waves in air is 3×10^8 m/s. Calculate the frequency of medium-wave transmitters which transmit waves of 300 m wavelength.
- 3 Water waves generated by a 5 Hz vibrator travel 4 m in 16 s. Find their speed and wavelength.
- 4 A fisherman's float bobs up and down nine times each minute as waves of wavelength 20 m pass his anchored boat. Find the speed of the waves.
- 5 Compression waves in water travel at 1450 m/s. A ship's depth-sounding device times the echo of such a pulse from the ship to the sea-bed and back. It takes 4 ms. How deep is the sea at this point?
- 6 Fig. 14.14 shows the position of a straight wave AB approaching a reflecting barrier. Copy the figure, and construct the new position of the wavefront when A reaches the barrier.

fig 14.14



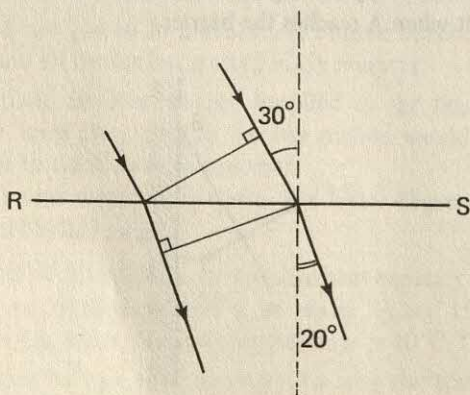
- 7 A small dipper in a ripple tank is 2 cm from a straight reflecting wall. The dipper starts to vibrate at 10 Hz creating circular water waves which travel at 10 cm/s. Make a diagram to show the positions of the waves (a) 0.2 s and (b) 0.3 s after the vibrator starts to operate.
- 8 Waves of wavelength 0.3 cm approach a boundary PQ in Fig. 14.15 and cross it. Two wavefronts are shown in the figure. Calculate the wavelength of the waves after they cross the boundary.

fig 14.15



- 9 After crossing the boundary RS in Fig. 14.16, waves of wavelength 1.0 cm travel at 3 cm/s. Calculate their speed and wavelength before crossing the boundary.

fig 14.16

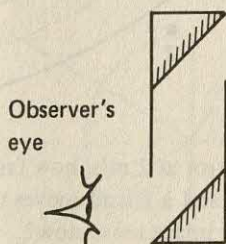


- 10 In an experiment with Young's slits using monochromatic light of wavelength 6×10^{-7} m, the double slits are 0.4 mm apart. Find the separation of the fringes observed in a plane 0.5 m from the slits.
- 11 Explain what differences there would be in the fringe pattern observed with Young's slits, if
- light of longer wavelength were used,
 - the double slits were set wider apart,
 - the fringes were observed in a plane twice as far from the slits.

CHAPTER 15

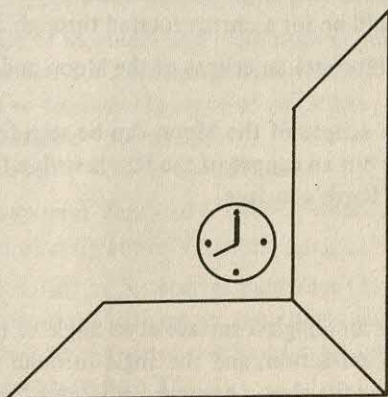
- 1 Fig. 15.12 shows a periscope for looking over the heads of crowds. Copy the figure and include a ray from the tip of an object to the observer's eye. Label the angle of incidence at each mirror.

fig 15.12



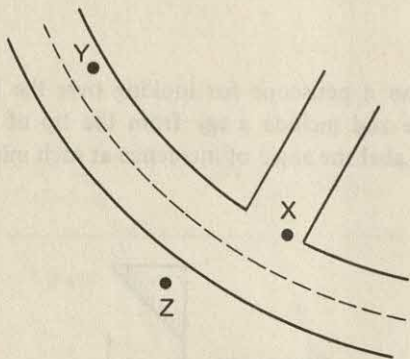
- 2 Sign your name normally; then repeat it in mirror writing, i.e. laterally inverted. Examine both signatures in a mirror.

fig 15.13



- 3 In Fig. 15.13 a vertical and a horizontal mirror are shown near a clock face at eight o'clock. Redraw the figure and include the images of the clock face seen in each mirror.
- 4 Fig. 15.14 shows a T-junction on a bend in a main road. A motorist at X can only see if the road is clear at Y by looking into the mirror at Z. Copy the figure and construct the mirror line.

fig 15.14



- 5 If you walk towards a mirror at 2 m/s, how fast does your image move? If you remain stationary and a friend moves the mirror towards you at 2 m/s, how fast does your image move now?
- 6 What is the shortest length of mirror in which a man 1.8 m tall can see the whole of his image while he is standing upright?
- 7 Construct a figure showing the reflection of a ray at a plane mirror. Then show the same ray being reflected after the mirror has been rotated through 20° . Measure the angle between the reflected rays. Deduce what this angle would be for a mirror rotated through 30° .
- 8 Draw diagrams to illustrate an eclipse of the Moon and an eclipse of the Sun.

Explain why an eclipse of the Moon can be seen from about half of the Earth's surface but an eclipse of the Sun is visible from only a much smaller area of the Earth's surface.

CHAPTER 16

- 1 A ray in air meets an air-glass surface at an angle of incidence of 60° . Find the angle of refraction, and the angle through which the ray is deviated on entering the glass. (Assume $n_{\text{air}}/n_{\text{glass}} = 1.5$.)

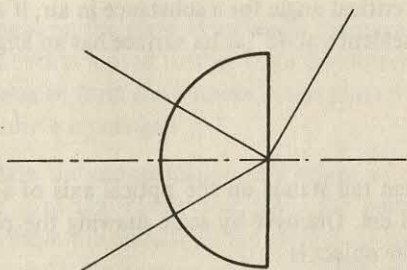
- 2 The speed of light in air is 3×10^8 m/s. In medium X its speed is 2×10^8 m/s, and in medium Y the speed of light is 2.5×10^8 m/s. Calculate

- (a) $_{\text{air}}n_X$,
- (b) $_{\text{air}}n_Y$,
- (c) $_Xn_Y$.

Which medium is optically most dense?

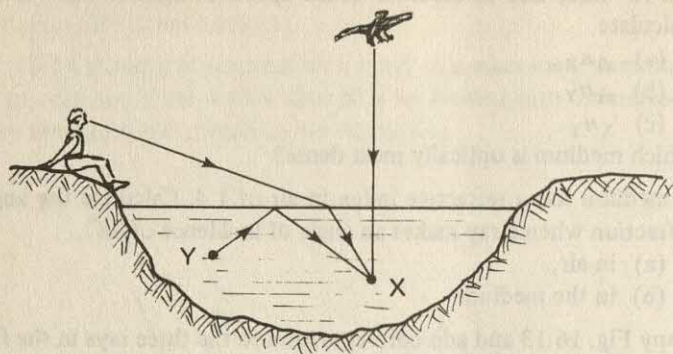
- 3 A medium has a refractive index in air of 1.4. Calculate the angle of refraction when a ray makes an angle of incidence of 28° ,
- (a) in air,
 - (b) in the medium.
- 4 Copy Fig. 16.13 and add correct arrows to the three rays in the figure. Explain how you decide which way the arrows should point.

fig 16.13



- 5 Calculate the critical angle for diamond using the graph in Fig. 16.3b.
- 6 Write down approximately the angle of refraction in water of refractive index 1.33, for a ray making an angle of incidence in air of 4.8° .
- 7 If i is the angle of incidence and r the angle of refraction, plot a graph of $\sin i$ against $\sin r$ for a substance whose refractive index in air is 1.4. Use the graph to discover the angle of refraction
- (a) for an angle $i = 30^\circ$ in air,
 - (b) for an angle $i = 30^\circ$ in the substance.
- 8 What is the apparent depth of a tank of water 24 cm deep, when it is viewed from directly above? (Assume $_{\text{air}}n_{\text{water}} = 1.33$.)
- 9 In Fig. 16.14 a fish at X receives light rays (a) from a bird directly overhead, (b) from the man on the river bank, and (c) from the fish at Y. What effect does the water's surface have on the light from each of these objects as it travels to X?

fig 16.14



- 10 Write out the second law of refraction and explain why the phrase '*of one colour*' is necessary.
- 11 Calculate the critical angle for a substance in air, if a ray in air making an angle of incidence at 48° at its surface has an angle of refraction of 30° .

CHAPTER 17

- 1 An object 2 cm tall stands on the optical axis of a converging lens of focal length 8 cm. Discover by scale drawing the position and size of the image if the object is
 - (a) 12 cm from the lens,
 - (b) 6 cm from the lens.
- 2 Define the terms 'optical centre', 'principal focus' and 'focal length' of a converging lens, and the 'magnification' of the image.
- 3 An image is cast on a screen 60 cm from a bright object by means of a single lens. If the magnification is 3, where must the lens be placed? Find its focal length by scale drawing.
- 4 Construct diagrams to illustrate the formation of a virtual image using
 - (a) a converging lens,
 - (b) a diverging lens.
 What is the chief difference between the two images?
- 5 At what distance from a converging lens of focal length 12 cm must an object be placed in order that an image of magnification 1 will be produced?
- 6 Describe a method for finding the focal length of
 - (a) a convex lens,
 - (b) a concave mirror.

- 7 Define 'radius of curvature' of a concave mirror, and describe how you would find it experimentally.
- 8 Make ray diagrams to illustrate the formation of
 - (a) a real image by a concave mirror,
 - (b) a virtual image by a concave mirror.
- 9 An object is 24 cm from a concave mirror, and its image is 16 cm from the mirror. Find the focal length and radius of curvature of the mirror, and the magnification of the image.
- 10 An object 20 cm from a spherical mirror gives rise to a virtual image 15 cm behind the mirror. Determine the magnification of the image, and the type of mirror used. Construct a diagram to scale in order to find the focal length of the mirror. Then calculate its radius of curvature and its power.

CHAPTER 18

- 1 How is the image on the plate of a pinhole camera affected if:
 - (a) the camera is moved further from the object,
 - (b) the distance from the pinhole to the plate is increased,
 - (c) the pinhole is enlarged?
- 2
 - (a) Calculate the magnification and height of the image when an object 1 m tall stands 5 m from the pinhole camera whose length from the pinhole to the plate is 10 cm.
 - (b) What would the answers to (a) be if the pinhole were replaced by a converging lens correctly focused?
- 3 What camera adjustment is necessary after photographing an object close at hand, before a distant scene can be filmed? Explain any change that is needed.
- 4 A photograph is taken with an exposure of $1/60$ s, and another is required to show the same scene with greater depth of field. Exposures of $1/125$ and $1/30$ s are possible, and the aperture can be made larger or smaller. What changes in the camera settings would you make?
- 5 An eye has a near point distance of 50 cm. What sort of lens in glasses would be needed to reduce the near point distance to 25 cm? Is this eye long-sighted or short-sighted?
- 6 Compare the methods of focusing in the camera and the human eye. An eye can see nothing in sharp focus which is further away than 20 m. What type of lens would help this eye to see distant objects?
- 7 Explain the functions of the following parts of a slide projector:
 - (a) projection lens,

- (b) condenser lens,
 - (c) concave mirror.
- 8 A slide projector projects an image magnified $\times 100$ onto a screen 10 m away.
 - (a) What is the distance between the slide and the projection lens?
 - (b) How does this distance compare with the focal length of the projection lens?
 - 9 Construct a ray diagram showing the functions of each lens in an astronomical telescope. How can the real image of the Moon's surface inside the telescope:
 - (a) be made larger,
 - (b) be made brighter?
 - 10 Describe a radio telescope, and outline some of its uses.
 - 11 Draw a ray diagram showing the optical system of a Newtonian reflecting telescope in normal adjustment. What advantages does a reflecting telescope have over a refracting telescope?
 - 12 Draw a ray diagram of a compound microscope. If the eyepiece magnifies $\times 10$, explain how a total magnification $\times 150$ is possible.

CHAPTER 19

- 1 Explain how a pure spectrum of white light can be produced on a screen. How would the appearance of the spectrum be changed if a sheet of yellow glass were placed in the path of the rays?
- 2 Beams of red and blue light overlap where they fall on a white screen. What is the resulting colour of the screen?
- 3 What colour results when
 - (a) magenta and yellow paints are mixed,
 - (b) yellow and peacock blue paints are mixed?
- 4 A red tomato rests on a yellow table cloth. What colours will the tomato and cloth appear in
 - (a) red light,
 - (b) green light?
- 5 What is meant by the electromagnetic spectrum? What evidence is there that both radio waves and light are electromagnetic waves?
- 6 Given that light travels at 3×10^8 m/s in free space and in air, find the frequency of
 - (a) radio waves of wavelength 240 m,
 - (b) γ rays of wavelength 10^{-14} m.

- 7 A sunbather inside a greenhouse finds it difficult to get a suntan. Explain the difficulty.
- 8 Which waves in the electromagnetic spectrum
 - (a) are usually associated with heat,
 - (b) have the longest wavelength?
 Name two properties common to all electromagnetic waves.

CHAPTER 20

(Where needed, take the velocity of sound in air as 330 m/s.)

- 1 Sketch the CRO waveform of a note produced by
 - (a) a tuning fork,
 - (b) another tuning fork giving a note of the same frequency but louder,
 - (c) a third tuning fork of the same loudness as (b) but with twice the frequency.
- 2 What is the wavelength in air of a note of frequency
 - (a) 330 Hz,
 - (b) 660 Hz?
- 3 Describe an experiment to determine the velocity of sound in air. What are the main sources of error in the experiment you have described?
- 4 A bat flying towards a wall emits a sound and receives an echo 30 ms later. How far is the bat from the wall?
- 5 What factors determine the pitch of a note emitted by a stretched string? Describe experiments you could perform to verify your statements. How can two strings emit notes of the same pitch and loudness but of different quality?
- 6 A tuning fork is held above a tall jar of water and the water is siphoned out slowly. At a certain point a loud note is heard. Explain why the loud note is heard and also how the experiment can be used to determine the velocity of sound.
- 7 A pipe closed at one end resonates when a tuning fork is held above it and the length of the air column is 33 cm. The next position of resonance is when the length of the air column is 99 cm.
 - (a) Calculate the wavelength of the note emitted by the tuning fork.
 - (b) Calculate the frequency of the note emitted by the tuning fork.

- 8 A stretched string is set vibrating by
- plucking the midpoint,
 - plucking it at a point one-quarter of its length from one end.

What is the ratio of the frequencies of the principal notes emitted?

A string has a tension of 20 N and its length is 100 cm. It emits a note of frequency 300 Hz. What note would be emitted if the tension were changed to 10 N and the length changed to 50 cm?

CHAPTER 21

- You are given a piece of thread and three bars, one of which is a magnet, the second of which is a magnetic material that is not magnetised, and the third of which is a non-magnetic material. Explain how you would distinguish between the three bars.
- Explain why hammering and heating tend to destroy magnetism.
- Lucy says 'N-poles attract because the N-pole of a magnet points towards the North Pole of the Earth'. Explain why she is wrong.
- What is meant by magnetic induction? Describe one experiment to illustrate magnetic induction.
- A test-tube has lots of iron filings in it and the open end is closed with a cork. The tube is shaken and then each end is brought up in turn to the N-pole of a suspended magnet. It attracts both ends of the magnet. The test-tube is then stroked with a magnet as shown in Fig. 21.3. One end now attracts the N-pole of the suspended magnet. Explain why and say which end of the test-tube it will be.
- Explain the terms 'magnetic meridian', 'angle of declination', and 'angle of dip'. A weak magnet is placed on a horizontal table with its N-pole pointing North. Explain how you would plot the lines of force around the magnet and sketch the pattern you would expect to find.
- Explain what is meant by the theory of tiny elementary magnets. Use your theory to explain
 - why there are no poles at the centre of a magnetised watch spring, but that if the watch spring is broken in half, poles appear at the two broken ends,
 - why there is a limit to the extent to which a steel rod can be magnetised.
- You are provided with two bar magnets and an unmagnetised rod of steel. Describe how you could magnetise the steel so that it had N-poles at each end.

- 9 A piece of watch spring is magnetised and then immersed in iron filings. Draw a diagram showing the appearance when it is withdrawn. It is then broken in half and again immersed in iron filings. Draw another diagram showing the appearance of each half.
- 10 What is meant by a neutral point. How would you demonstrate the existence of neutral points? How would the position of a neutral point between two bar magnets placed apart with their axes parallel and N-poles facing each other help you to decide which was the stronger magnet?

CHAPTER 22

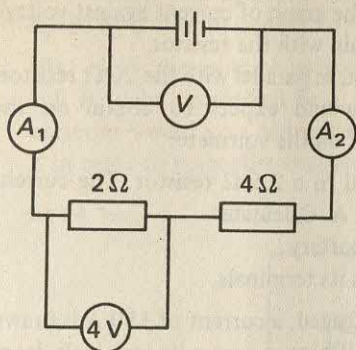
- 1 Explain why small pieces of paper are attracted to charged objects even though the paper itself is not electrically charged.
- 2 Why is it not possible to charge a conductor by holding it in your hand and rubbing it with a duster?
- 3 How many electrons would it be necessary to remove from an object in order to give it an overall charge of 6×10^{-8} C? (Charge on one electron = 1.6×10^{-19} C.)
- 4 A gold-leaf electroscope is given a negative charge. When an object of unknown charge is placed near the *bottom* of the rod, the gold leaf is observed to collapse slightly. State, with explanation, whether you think the object has a positive or negative charge.
- 5 Two objects, A and B, are known to be positively charged. How would you determine which of them had the larger charge?
- 6 Explain whether or not it would be possible to charge two sheets of plastic in the manner described in Section 22.6.
- 7 Charged objects do not retain their charge indefinitely. Explain how they become discharged.
- 8 Two magnets can *both* attract and repel each other whereas two charged objects will *either* attract *or* they will repel each other. Account for this difference in behaviour.
- 9 An alternating voltage of maximum value 10 V has a frequency of 100 Hz. It is applied to the input terminals of a cathode ray oscilloscope. Draw carefully the pattern you would expect to observe on the oscilloscope if the 'y sensitivity' control was set at 5 V/cm and the time base at 10 ms/cm. (Include in your diagram the horizontal and vertical lines on the screen of the oscilloscope which you can assume are 1 cm apart.)

- 10 What is the purpose of the following features in a cathode ray oscilloscope:
- the filament,
 - the cathode,
 - the anode,
 - the vertical y deflection plates,
 - the fluorescent screen?

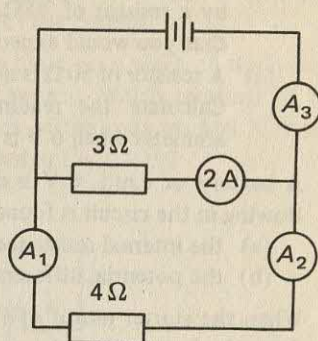
CHAPTER 23

- Calculate the values you would expect to observe on the ammeters and voltmeters in the circuits in Fig. 23.23.
- 20 kJ of heat are produced in a heater when 2 A flows for 100 s. What is
 - the charge that flows,
 - the potential difference across the heater,
 - the resistance of the heater?
- Explain why the resistance of an ammeter should have a low value, whereas that of a voltmeter should have a high value.
- A series circuit consists of three resistors of value $2\ \Omega$, $4\ \Omega$, and $6\ \Omega$, connected to a 20 V battery.
 - How does the potential difference across the $6\ \Omega$ resistor compare with that across the $2\ \Omega$ resistor?
 - How does the current in the $6\ \Omega$ resistor compare with that in the $2\ \Omega$ resistor?
 - Calculate the value of the potential difference across the $4\ \Omega$ resistor.
 - Use your answer to (c) to calculate the potential difference across the $2\ \Omega$ and $6\ \Omega$ resistors.
- A circuit consists of three resistors of value $2\ \Omega$, $4\ \Omega$ and $6\ \Omega$ connected in parallel to a 20 V battery.
 - How does the potential difference across the $6\ \Omega$ resistor compare with that across the $2\ \Omega$ resistor?
 - How does the current in the $6\ \Omega$ resistor compare with that in the $2\ \Omega$ resistor?
 - Calculate the value of the current in the $4\ \Omega$ resistor.
 - Calculate the total current drawn from the battery.
- Draw a diagram of a circuit which could be used to obtain the values of current and voltage required to plot the graph shown in Fig. 23.12.

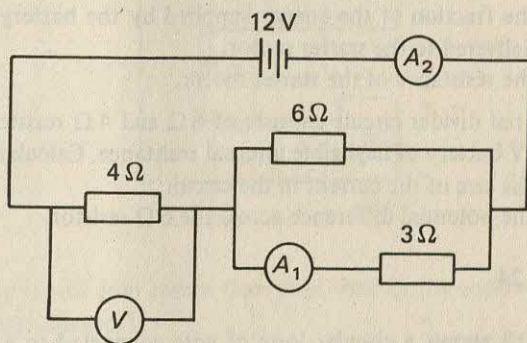
fig 23.23



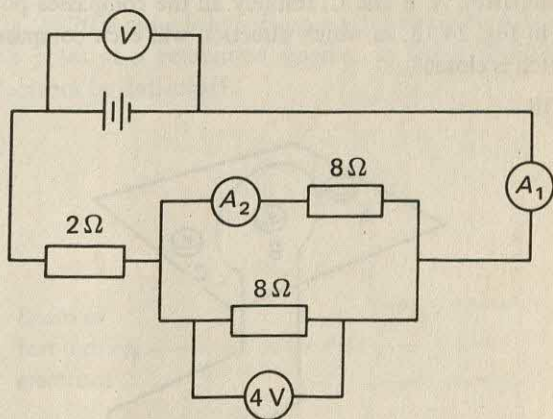
(a)



(b)



(c)



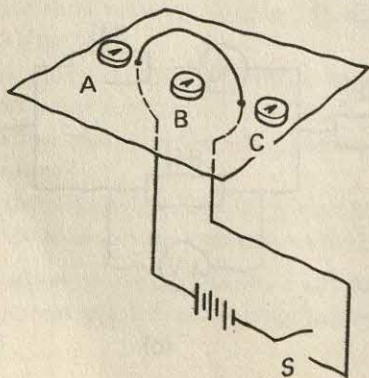
(d)

- (b) The filament lamp is now removed from the circuit and replaced by a resistor of $20\ \Omega$. Draw the graph of current against voltage that you would expect to obtain with the resistor.
- (c) A resistor of $50\ \Omega$ is now placed in parallel with the $20\ \Omega$ resistor. Calculate the reading you would expect to obtain on the ammeter when $6\ \text{V}$ is recorded on the voltmeter.
- 7 A battery of e.m.f. $6\ \text{V}$ is connected to a $2.5\ \Omega$ resistor. The current flowing in the circuit is found to be $2\ \text{A}$. Calculate:
- the internal resistance of the battery,
 - the potential difference across its terminals.
- 8 When the starter motor of a car is engaged, a current of $150\ \text{A}$ is drawn from the battery and the potential difference across its terminals drops from $12\ \text{V}$ to $9\ \text{V}$. Calculate
- the internal resistance of the battery,
 - the fraction of the energy supplied by the battery which is not delivered to the starter motor,
 - the resistance of the starter motor.
- 9 A potential divider circuit consists of $6\ \Omega$ and $4\ \Omega$ resistors connected to a $1.5\ \text{V}$ battery of negligible internal resistance. Calculate
- the size of the current in the circuit,
 - the potential difference across the $6\ \Omega$ resistor.

CHAPTER 24

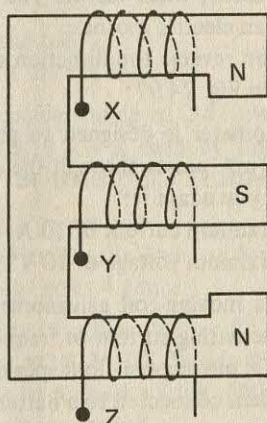
- 1 Fig. 24.18 shows a circular loop of wire connected to a battery via a switch. The coil passes through a card on which are placed three plotting compasses, A, B and C. Initially all the compasses point North as shown in Fig. 24.18. In which direction will each compass point when the switch is closed?

fig 24.18



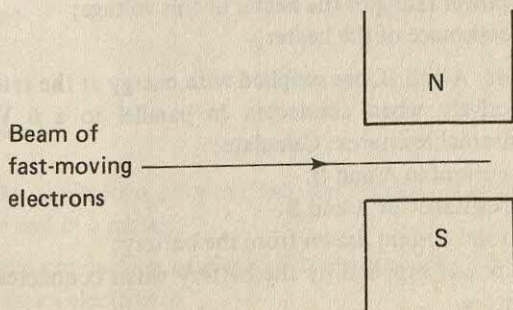
- 2 How would you establish that the strength of the magnetic field produced by a current depends on the size of the current?
- 3 Fig. 24.19 shows three coils of wire wound on a piece of steel. Each coil is to be connected to a separate battery. Which terminal of the batteries would you connect to the ends of the coils marked X, Y and Z in order to magnetise the steel as shown in Fig. 24.19?

fig 24.19



- 4 (a) Why is soft iron rather than steel used in the construction of an electromagnet?
(b) If steel were used by mistake in the electromagnet of an electric bell, how would this affect the operation of the bell?
- 5 (a) Fig. 24.20 shows a beam of fast-moving electrons passing through the poles of a permanent magnet. In which direction will the electrons be deflected?

fig 24.20



- (b) Would a stationary electron experience any force when placed between the poles of a magnet?
- 6 Explain the purpose of the following features of an electric motor:
- the coil,
 - the magnetic field,
 - the brushes,
 - the split-ring commutator.
- 7
- State, with reasons, what factors you think control the speed of rotation of an electric motor.
 - How would you reverse the direction of rotation of the electric motor shown in Fig. 24.9?
- 8 A moving-coil galvanometer is designed to give a full-scale deflection when a current of 50 mA passes through the coil, and has a resistance of $0.2\ \Omega$. How would you adapt it:
- to record a maximum current of 10 A,
 - to record a maximum voltage of 10 V?
- 9
- Explain why a moving-coil galvanometer could not be used to measure an alternating current of frequency 50 Hz.
 - Explain why a moving-coil loudspeaker would not produce a sound wave when connected to a battery.

CHAPTER 25

- 1 The current in a small electric heater is 3.5 A when it is connected to a 12 V battery. Assuming the internal resistance of the battery is negligible, calculate:
- the number of joules liberated as heat energy by each coulomb;
 - the number of coulombs that pass through the heater in one second;
 - the power rating of the heater at this voltage;
 - the resistance of the heater.
- 2 Two resistors, A and B, are supplied with energy at the rate of 4 W and 6 W respectively when connected in parallel to a 6 V battery of negligible internal resistance. Calculate:
- the current in A and B;
 - the resistance of A and B;
 - the total current drawn from the battery;
 - the power supplied by the battery when connected to the two resistors.

- 3 An electric heater is found to liberate 1080 J of energy in one minute when it is connected to a 6 V battery of negligible internal resistance. Calculate:
- (a) the current in the heater;
 - (b) the resistance of the heater;
 - (c) the power of the heater at this voltage.
- 4 A 1 kW electric fire is connected to the 240 V alternating mains supply. Calculate:
- (a) the current in the coil;
 - (b) the resistance of the heating coil.
- 5 A tumble dryer has a power rating of 2.43 kW when connected to the 240 V mains supply. 0.18 kW is used by the electric motor, the remainder is supplied to the heater coil. Calculate:
- (a) the current in the heater coil;
 - (b) the total cost of leaving the dryer on for 45 min.
- (Assume 1 kWh of energy costs 4 pence.)
- 6
- (a) Calculate the size of a fuse you would fit in the three-pin plug used to connect a 240 V, 700 W hair dryer to the mains supply.
 - (b) Would it be safe to connect a 3 kW, 240 V electric fire to a 13 A socket? Explain your answer.
 - (c) An electric cooker has four heating elements each rated at 3.2 kW when connected to a 240 V supply. Calculate the total current drawn from the mains supply when all four elements are used at the same time.
- 7 Mains adaptors can be bought which enable more than one appliance to be connected to a single 13 A socket. What precautions must be taken when using such adaptors?
- 8 The electricity bill for a particular household comes to £45 for a three-month period. Assuming that 1 kWh costs 4 pence, calculate the average daily energy consumption over this period. (Take three months as 90 days.)

CHAPTER 26

- 1 Explain the difference between the conduction of electricity in an electrolyte and in a metal.
- 2 Compare the conduction of electricity in a gas with the conduction of electricity in an electrolyte.

- 3 Explain the difference between:
 - (a) an ion and an atom;
 - (b) a positive ion and a negative ion;
 - (c) an electrolyte and a non-electrolyte.
- 4 Describe an experiment you would perform to verify Faraday's law of electrolysis. State clearly what readings you would take and how you would use them to confirm the truth of the law.
- 5 Draw a carefully labelled diagram of a voltameter that could be used to plate a nickel spoon with a thin coating of silver. State, with reasons, how you would connect a battery to the voltameter in order to plate the spoon.
- 6 A voltameter consisting of two silver electrodes immersed in silver nitrate solution is connected in series with another voltameter consisting of two copper electrodes immersed in copper sulphate solution. The silver nitrate electrolyte contains Ag^+ ions and the copper sulphate electrolyte contains Cu^{2+} ions. A current of 2 A is passed through both voltameters for five minutes.
 - (a) Calculate the total amount of electric charge that passes through the voltameters.
 - (b) Explain why the number of silver *atoms* liberated during the electrolysis is double the number of copper atoms.
 - (c) How many copper *atoms* are produced at the cathode of the copper voltameter? (The charge on one electron is $1.6 \times 10^{-19} \text{ C}$.)
- 7 It is found that either 2.89×10^5 or $1.93 \times 10^5 \text{ C}$ of electric charge can liberate 6×10^{23} atoms (1 mole) of iron atoms. Account for this difference in terms of the charge on the iron ions. (The charge on one electron equals $1.6 \times 10^{-19} \text{ C}$.)
- 8 It is found that 2.7 g of nickel are liberated at the cathode of a nickel voltameter when a current of 1.25 A flows through the voltameter for 2 h. Calculate the electrochemical equivalent of nickel.
- 9
 - (a) Explain the difference between a primary cell and a secondary cell.
 - (b) Why do secondary cells eventually have to be replaced?
 - (c) Describe in detail the construction of a dry cell.

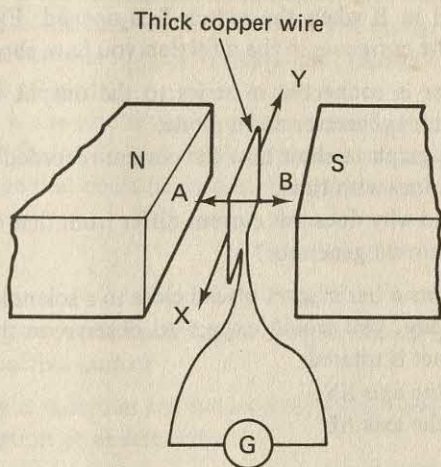
CHAPTER 27

- 1 The speed of rotation of the coil in an alternating-current generator is increased. Explain what changes, if any, will occur in the amplitude and frequency of the current produced by the generator as a result of this change.

- 2 Fig. 27.14 shows a piece of thick copper wire, which has been bent into the form of a square, and is held between the poles of a permanent magnet. Explain what deflection you would expect to observe on the galvanometer G when the wire is moved backwards and forwards:

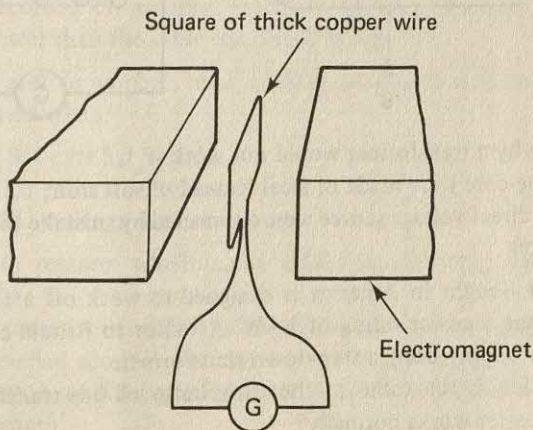
- in the direction AB,
- in the direction XY.

fig 27.14



- 3 A square frame of thick copper wire is placed between the poles of an electromagnet as shown in Fig. 27.15 and any induced current in the

fig 27.15

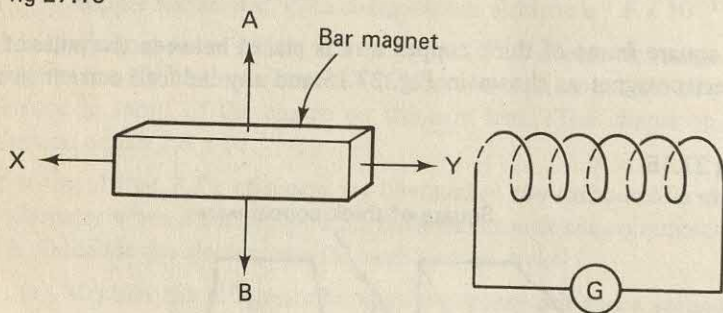


coil is recorded on the galvanometer G. What deflection, if any, would you expect to observe on the galvanometer when the electromagnet is

- (a) switched on,
- (b) left on for a short time,
- (c) switched off?

- 4 (a) Why is there no current recorded on the galvanometer G in Fig. 27.8 when the current in solenoid A is steady?
- (b) Copy Fig. 27.8 and on it indicate the direction of the current induced in B when the switch S is opened. Explain why you think the current is in the direction you have shown.
- 5 A galvanometer is connected in series to the output terminals of an alternating-current generator and a diode.
 - (a) Draw a graph to show how the current recorded on the galvanometer varies with time.
 - (b) How and why does this current differ from that obtained from a direct-current generator?
- 6 Fig. 27.16 shows a bar magnet placed close to a solenoid. Explain what deflection, if any, you would expect to observe on the galvanometer when the magnet is rotated
 - (a) about the axis XY,
 - (b) about the axis AB.

fig 27.16



- 7 Explain why a transformer would not work if
 - (a) the core were made of steel instead of soft iron;
 - (b) a direct voltage source were connected by mistake to the primary coil.
- 8 A toaster bought in America is designed to work off a 120 V supply when it has a power rating of 1 kW. A visitor to Britain connects it to the 240 V supply using a step-down transformer.
 - (a) What is the value of the turns ratio of this transformer if the toaster works normally?

- (b) What is the size of the current in the secondary coil of the transformer?
 - (c) Assuming that the transformer is 100% efficient what is the size of the current in the primary coil?
- 9 The voltage developed by the generator at a power station is 25 000 V. A transformer is used to step up this voltage to 400 000 V for transmission through the grid system.
- (a) What is the value of the turns ratio of this transformer?
 - (b) What is the ratio of the current in the primary coil of the transformer to that in the secondary coil? (Assume the transformer is 100% efficient.)
 - (c) Why is the electrical energy transmitted at such a high voltage?
 - (d) What factor do you think controls the maximum value of the voltage that could be used in the grid system?

CHAPTER 28

- 1 How would you determine the type or types of radiations being emitted from a radioactive source?
- 2 Explain why α radiation can be detected using a spark counter whereas β and γ radiation go undetected.
- 3 Explain why it is necessary to use a GM tube with a thin mica window in order to detect α radiation.
- 4
 - (a) Most radioactive sources used in schools have an activity of 5 μCi . How many ionising radiations would be emitted from such a source in one second?
 - (b) A GM tube is placed close to the source and the count rate is recorded on a scaler. Give two reasons why this count rate is lower than the value calculated in (a).
- 5 Give the atomic number, mass number and chemical symbol of the isotope formed when:
 - (a) ${}_{38}^{90}\text{Sr}$ emits a β particle,
 - (b) ${}_{27}^{60}\text{Co}$ emits γ radiation,
 - (c) ${}_{92}^{234}\text{U}$ emits an α particle.
- 6 State, with reasons, whether you think the following statements are true or false:
 - (a) α particles have lost all their kinetic energy after they have travelled about 7 cm in air.
 - (b) Both α and β particles would have an infinitely long range in vacuum.

- 7 A radioactive isotope has a half-life of 20 min. A source of this isotope is found to have an activity of $50 \mu\text{Ci}$. What would you expect the activity to be after:
- (a) 20 min,
 - (b) 40 min,
 - (c) one hour?
- 8 A source of α radiation lying on a bench is *less* dangerous than one emitting γ radiation. However, if a radioactive gas emitting α radiation is inhaled it is *more* dangerous than one which emitted γ radiation. Explain why both these statements are correct.

ANSWERS TO QUESTIONS

CHAPTER 1

1. 8.9 g/cm^3 2. 8920 kg/m^3 3. $3.70 \times 10^{-3} \text{ m}^3$ 4. 2.3 mm
5. 1.29 kg/m^3 6. 87 kg 7. (a) 1.79 cm (b) 30.8 g (c) 1.60 cm
8. (a) B heaviest, D lightest (b) C smallest, A and B largest (c) E

CHAPTER 2

1. (a) 5 Pa (b) 6 Pa (c) 2000 N (d) $6.0 \times 10^5 \text{ N}$
2. (a) 25 kPa , 6.25 kPa 3. 3.75 kPa 6. (a) 49.5 kPa , 99 kPa , 10^4 kPa
(b) 214 kPa , 428 kPa 7. (a) Y (b) Both rise together (c) 50 N
(d) 2.0 N , 8.0 N , 12. $1.22 \times 10^4 \text{ Pa}$, 3.85 N
15. 4 m^2

CHAPTER 3

1. (a) 5 N at 53.1° to 3 N force (b) 26.5 N at 40.8° to 10 N (c) 42.0 N
at 30.4° to 15 N force 2. 204 km/h at 78.8° to wind 3. 825 N at
 14.0° ahead of vertical 4. 0.58 N , 1.15 N
5. 77300 N on bisector of force directions 7. 5.83 km/h in direction
downstream at 59° to bank 8. 6.0 N at 83.6° to 3 N direction
9. 115 N

CHAPTER 4

2. 5 cm/s , 32 cm/s , 3. 50 cm/s 4. 83.3 m/s^2 5. 500 cm/s^2
6. 5 m/s^2 7. 100 N 9. (a) 2.5 m/s^2 (b) 1.25 m (c) 500 m
10. (a) 3.0 N (b) 9.80 m/s^2 , 9.77 m/s^2 11. 1.67 m/s in same direction
as 2 kg initially moving 12. 0.67 m/s in original direction of motion
13. 0.50 m/s 14. 6.3 m/s

CHAPTER 5

3. (a) 30 N (b) 2.5 N (c) 1.75 N (d) 2.32 N. 4. 1.33 m 5. 2.4 m on opposite side to Jim, 2.33 m on same side as Mary. 6. (a) zero (b) 24 Nm 7. 2.5 N 8. (a) 1.65×10^4 N (b) 1.56×10^5 N and 1.74×10^5 N

CHAPTER 6

1. (a) 12 J (b) 100 J 2. (a) 200 J (b) 1.0 J 3. (a) 1000 J (b) 10 000 J
4. 5 000 MW 5. (a) 3.47×10^4 J (b) 2.7×10^8 J (c) 8.0×10^9 J
6. 0.67 m/s^2 , 33.3 N 8. (a) 12 000 J (b) 1200 J/s (c) 1200 W
9. (a) 1.2×10^6 J (b) 1.2×10^6 J, 4000 W

CHAPTER 7

1. 80% 2. (a) 1500 J (b) 1.25 m (c) 1250 J (d) 0.83 3. (a) 1500 J (b) 156 N (c) 1875 J 4. (a) 942 (b) 53% (c) 942 J 5. 600 J, 600 J, 1000 J, 20 N
6. 0.25 8. (a) 2.5 (b) 160 N (c) 4000 J

CHAPTER 8

2. (a) 80 g (b) 0.80 N (c) 0.70 N (d) 0.72 N 3. (a) 0.30 N (b) 30 g (c) 30 cm^3 (d) 30 cm^3 (e) 30 cm^3 (f) 0.24 N (g) 24 g (h) 0.80 g/cm^3 (i) 1.2 g/cm^3
4. (a) 0.60 g/cm^3 (b) 6.0 N (c) 6.0 N (d) 600 cm^3 (e) 4.0 N
5. 13.0 N, 12.1 N 8. 2.86 m^3 9. 5 cm, 6.25 cm.

CHAPTER 9

1. (a) 2200 J/kgK (b) 4.3°K 2. 300 J/kg K 3. 1000 J, 0.25°K
4. 6000 J, 1.4 K 5. 5 saucepans (4.6) 6. 725°C 7. (a) 1.6×10^{14} J (b) 6.15×10^4 kg/s (c) 9.4 days 8. (a) 10 000 J/min (b) 6.0×10^7 J (c) 6.0×10^{11} J (d) 8.3 MW

CHAPTER 10

1. (a) 1.5 mm (b) 60 mm 4. 2.0×10^7 per K 6. 20.2°C

CHAPTER 11

6. $1.67 \times 10^4 \text{ N/m}^2$ 7. 2.25 atmospheres 8. $1.53 \times 10^5 \text{ N/m}^2$
9. 0.48 g/dm^3 10. 87°C

CHAPTER 13

2. 2.3×10^7 (b) 2.1×10^5 J 3. (a) 2.3×10^5 J (b) 4.2×10^4 J
 (c) 3.3×10^4 J (d) 3.1×10^5 J 4. 18.3 min 5. 2.1 mins, 0.24 kg
 6. 92 000 J 7. (a) 11 kJ/min (b) 7 m 38 s (c) 41 min 49 s.
 8. (a) 40 J (b) 34.5 kJ (c) 38.28 kJ (d) 23.05 g

CHAPTER 14

1. (c) 30 Hz (d) 3.75 Hz 2. 1 MHz 3. 5 cm 4. 3 m/s 5. 2.9 m
 8. 0.2 cm 9. 0.68 cm, 2.05 cm/s 10. 7.5×10^{-4} m

CHAPTER 15

5. 2 m/s, 4 m/s

CHAPTER 16

1. 35.3° , 24.7° 2. (a) 1.5 (b) 1.2 (c) 0.8 3. (a) 19.6 (b) 41.4
 5. 24.8° 7. 20.9° , 44.4° 8. 18 cm 11. 42.3°

CHAPTER 17

1. (a) 24 cm, 4 cm (b) 24 cm, 8 cm. 3. 15 cm, 11.25 cm 5. 24 cm
 9. 9.6 cm, 19.2 cm, 0.67 10. 0.75, - 60 cm, - 120 cm, - 1.67 D

CHAPTER 18

2. (a) 0.02, 2 cm (b) the same 5. 2 D 8. (a) 10 cm (b) 9.9 cm

CHAPTER 19

6. (a) 1.25 MHz (b) 3×10^{22} Hz

CHAPTER 20

2. (a) 1 m (b) 0.5 m 7. (a) 1.32 m (b) 250 Hz

CHAPTER 22

3. 3.75×10^{11}

CHAPTER 23

- (a) $A_1 = A_2 = 2A$, V reads 12V (b) $A_1 = A_2 = 1.5 A$, $A_3 = 3.5 A$
(c) $V = 8$ volts, $A_2 = 2A$, $A_1 = 1\frac{1}{3} A$ (d) $A_2 = 0.5 A$, $A_1 = 1 A$,
 $V = 6$ volts
- (a) 200 C (b) 100 V (c) 50Ω 4. (a) $3x$ (b) same (c) $6\frac{2}{3} V$
(d) $3\frac{1}{3} V$, 10 V 5. (a) same (b) $\times \frac{1}{3}$ (c) 5 A (d) $18\frac{1}{3} A$
- (a) 0.5Ω (b) 5 V 8. (a) 0.02Ω (b) $\frac{1}{4}$ (c) 0.06Ω
- (a) 0.15 A (b) 0.9 V

CHAPTER 24

- (a) $1.005 \times 10^{-3} \Omega$ in parallel (b) 199.8Ω in series.

CHAPTER 25

- (a) 12 J (b) 3.5 C (c) 42 W (d) 3.4Ω 2. (a) $\frac{2}{3} A$, 1 A (b) 9Ω ,
 6Ω (c) $1\frac{2}{3} A$ (d) 10 W 3. (a) 3 A (b) 2Ω (c) 18 W
- (a) 4.16 A (b) 57.6Ω 5. (a) 9.4 A (b) 7.3 p 6. (a) 3 A (c) 53.3 A
g. $4.5 \times 10^7 J$ (12.5 kWh)

CHAPTER 26

- (a) 600 C (c) 1.9×10^{21} atoms 8. $3 \times 10^{-7} kg/C$

CHAPTER 27

- (a) 2:1 (b) 8.33 A (c) 4.16 A 9. (a) 1:16 (b) 16:1

CHAPTER 28

- (a) $1.85 \times 10^5/s$ 5. (a) 39,90 (b) 27,60 (c) 90,230
- (a) 25 μCi (b) 12.5 μCi (c) 6.25 μCi

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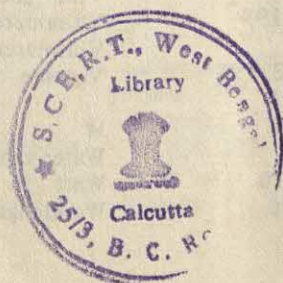
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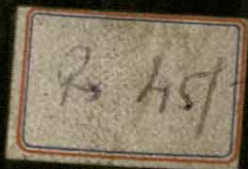
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